tions $P_{+}(\mathbf{x})$ and $P_{-}(\mathbf{x})$ in the domain $\Omega$, we can determine the topological picture of the distribution of the sets $H_{1}, H_{2}, H_{3}$ in $\Omega$, and, in accordance with the Theorems 1.1.1.4 and 1.5, we can classify the stability of the trivial solution of the system(1.2). Lemma 1.1 makes it possible to carry over this classification to the system (1.1).

The author thanks V. V. Rumiantsev for his interest to this paper.

## REFERENCES

1. Chetaev, N. G., Stability on Motion, Gostekhizdat, Moscow, 1955.
2. Malkin, I, G. Theory of the Stability of Motion. "Nauka", Moscow, 1966.
3. Matrosov, V. M. , On a theory of the stability of motion. PMM Vol, 26, $\mathrm{N}^{*} 6,1962$.
4. Nemytskii, V.V. and Stepanov, V.V., Qualitative Theory of Differential Equations. (English translation) Princeton Univ. Press, Princeton, N, J. 1960.
5. Nemytskii, V. V., On the problem of qualitative study in the large by means of Liapunov function methods. Vestn. Mosk. Gosudar. Univ., Ser. Matem., Mekhan, , $\mathrm{N}^{\mathrm{F}} 6,1962$.
6. Oleinik, O. A., On second order linear equations with a nonnegative characteristic form. Matem, $\mathrm{Sb}_{\mathrm{o}}, \mathrm{Vol}_{6} 69$, $\mathrm{N}^{2} 1,1966$.
7. Oleinik, O.A, and Radkevich, E. V., Second order equations with a nonnegative characteristic form. In: Matematicheskii Analiz, Itogi Nauki, VINITI Akad. Nauk SSSR, 1971.
8. Gantmakher, F. R. . Theory of Matrices. (English translation) Vols. 1, 2, Chelsea, New York, 1973.
9. Sobolev, S. L. . Applications of Functional Analysis in Mathematical Physics. (English translation), American Math. Soc. , Vol. 7., Providence, R. I. , 1963.
10. Samarskii, A. A., Introduction to the Theory of Difference Schemes. "Nauka". Moscow, 1971.

Translated by J. F.H.

UDC 531.36

# ASYMPTOTIC METHOD FOR MULTI-DIMENSIONAL SYSTEMS OF PARTIAL DIFFERENTLAL EQUATIONS 

PMM Vol, 39, $\mathrm{N}^{2} 4,1975$, pp. 668-675
A. A. ROZENBLIUM
(Gor ${ }^{\text {kii) }}$
(Received July 31, 1974)

We consider almost linear symmetric hyperbolic systems with constant coefficients in the linear part and with nonlinear terms containing a small parameter. The asymptotic method used here for construction of approximate solutions is based on the work of Bogoliubov and Mitropol'skii [1], and has been applied to systems with a single independent spatial variable [2,3]. Along with a slow time we introduce slow coordinates. For the approximate solution we obtain, not an infinite system as in [4], but a finite system of almost linear partial differential equations with constant coefficients, a system which is simpler than the original one. We present an algorithm for obtaining approximate solutions, We also show
that the approximate solution is close to the exact solution on an arbitrary finite interval.

1. We consider the system

$$
\begin{align*}
& u_{t}+\sum_{l=1}^{s} A^{(l)} u_{l}+B u=\mu f(x, t) P \nmid(u)  \tag{1.1}\\
& u(x, t)=\left\{u^{(1)}, \ldots, u^{(n)}\right\}, \quad x=\left\{x_{1}, \ldots, x_{s}\right\} \\
& u_{l}=\partial u / \partial x_{l} \\
& f(x, t)=\left\{f^{(1)}, \ldots, f^{(n)}\right\}, \quad f \in C_{r+3}, \quad r>s / 2+1 \\
& f P=\left\{f^{(1)} P^{(1)}, \ldots, f^{(n)} P^{(n)}\right\} \\
& f(x, t)=\sum_{m=-\infty}^{\infty} f_{m} \exp \left[i\left(m_{0} v t \mid-\sum_{l=1}^{s} m_{l} u_{l} x_{l}\right)\right] \\
& m=\left\{m_{0}, m_{1}, \ldots, m_{s}\right\}, \quad P=\left\{P^{(1)}, \ldots, P^{(n)}\right\}
\end{align*}
$$

Here $A^{(l)}$ and $B$ are constant real square matrices of order $n ; A^{(l)}$ are symmetric matrices: $\mu$ is a small parameter, and $P(u)$ is a polynomial in $u$.
Along with the system (1.1) we consider the following degenerate linear system with constant coefficients

$$
\begin{equation*}
u_{t}+\sum_{l=1}^{s} A^{(l)} u_{l}+B u=0 \tag{1.2}
\end{equation*}
$$

This system has solutions of the form:

$$
\begin{equation*}
u(x, t)=a \psi \exp [i(\omega t-k x)]+c_{0} c_{*}, \quad k=\left\{k_{1}, \ldots, k_{s}\right\} \tag{1,3}
\end{equation*}
$$

(the abbreviation $c, c$. indicates the complex conjugate of the expression appearing before it). Here, $a$ is an arbitrary complex number; $\psi$ is a right zero vector of the matrix $H ; \omega$ and $k$ are connected by means of the dispersion equation

$$
\begin{equation*}
D(\omega, k)=\operatorname{det} H=0, \quad H=\omega I-\sum_{l=1}^{s} A^{(l)} k_{l}-i B \tag{1.4}
\end{equation*}
$$

where $I$ is the unit matrix
Let there exist a nonempry set $\Omega$ of real frequencies $\omega_{c}$ mutually commensurable and also commensurable with $v$; let $\Omega$ also contain wave vectors $k_{c}=\left\{k_{\mathrm{c} 1}, \ldots\right.$, $\left.k_{c l}\right\}$ with real components, wherein like-named components $k_{1 s}, \ldots, k_{c l}, \ldots$ are mutually commensurable and are also commensurable with $x_{1}$. The wave vector components satisfy the dispersion equation $D\left(\omega_{c}, k_{c}\right)=0$. We assume also that

$$
\begin{equation*}
\partial D\left(\omega_{c}, k_{\varepsilon}\right) / \partial \omega \neq 0 \tag{1,5}
\end{equation*}
$$

Consequently, in neighborhoods of the points $\left(\omega_{c}, k_{c}\right) \mathrm{Eq}(1.4)$ defines $\omega$ as a singlevalued function of $k$. Further, we assume that the values $d \omega /\left.d k_{l}\right|_{k=k}{ }_{c l}$ are real. To each set of values of ( $\omega_{c}, k_{c}$ ) there corresponds an eigen-wave of the system (1.1) of the form (1.3).

We pose the problem of constructing an approximate solution of the system (1.1), generated by an initial combination of waves of the form

$$
\begin{equation*}
\sum_{c=1}^{d} a_{c}^{(0)} \lim _{c} \exp \left(-i k_{c} x\right)+c_{c} c_{v}, \quad-\infty<x<\infty, \quad\left(\omega_{c}, k_{c}\right) \in \Omega \tag{1,6}
\end{equation*}
$$

where $a_{c}{ }^{(0)}$ are specified complex numbers.
We assume the following conditions to be satisfied.

1. The set $N_{1}$ of pairs of vectors

$$
\begin{aligned}
& \left(m^{(b)}, n^{(b)}\right), \quad m^{(b)}=\left\{m_{0}^{(b)}, m_{1}^{(b)}, \ldots, m_{s}^{(b)}\right\} \\
& n^{(b)}=\left\{n_{1}^{(b)}, \ldots, n_{q^{(b)}}^{(b)}\right.
\end{aligned}
$$

with integral components, for which

$$
\begin{aligned}
& b=d+1, \ldots, q, \quad n^{(b)} \omega_{0}=\sum_{c=1}^{d} n_{c}^{(b)} \omega_{e}, \quad n^{(b)} h^{(0)}=\sum_{c=1}^{d} n_{s}^{(b)} k_{c t}
\end{aligned}
$$

is finite, $i_{s}$, the system has a finite number of combinative resonance waves ( $\omega_{c}$, $\left.k_{c}\right) E \Omega, c=d+1, \ldots, q$.
$2^{*}$. There exists a constant $L_{1}$ such that

$$
\begin{aligned}
& \inf \mid D\left(m_{0}^{(b)} v+n^{(b)} \omega, \quad m_{1}^{(b)}{h_{1}}^{(b)} n^{(b)} k^{(1)}, \ldots\right. \\
& \left.m_{s}^{(b)} x_{s}+n^{(b)} k^{(s)}\right) \mid=L_{1}>0_{t} \quad\left\{m^{(b)}, n^{(b)} \mid=N_{1}\right.
\end{aligned}
$$

This requirement means that the number of combinative waves arbitrarily close to resonance waves is finite. (in practical calculations one should regard combinative waves for which $D(\omega, k) \sim \mu$ as resonance waves).

Conditions $1^{\circ}$ and $2^{\circ}$ are satisfied. in particular, speaking geometrically, the surface in $E_{s+1}$ (the space of $\omega_{0}, k_{1}, \ldots, k_{s}$ ), defined by the dispersion equation, is bounded with respect to one of the coordinates.

We seek an approximate solution of the system (1, 1) in the half-space ( $-\infty<x<$ $\infty, 0<t<\infty)$, belonging to the class $C_{r}$, in the form

$$
\begin{align*}
& u^{(1)}(x, t)=\sum_{c=1}^{q} a_{c}(\chi, \tau) \psi_{c} \exp \left[i\left(\omega_{c} t-h_{c} x\right)\right]+  \tag{1.7}\\
& \quad c_{0} c_{0}+\mu w(x, t, \chi, \tau), \quad-\infty<x<\infty, \quad 0<t<\infty \\
& x=\chi_{0}+\mu x, \quad \tau=t_{0}+\mu t
\end{align*}
$$

Here, $\chi_{0}$ and $\tau_{0}$ are constants; $a_{c}(\chi, \tau)$ are unknown scalar complex-valued functions of the slow variables $\chi$ and $\tau$; $w(x, t, \chi, \tau)$ is an unkown real vector-valued function, periodic in 2 and $t$ with the periods $\Lambda=\left\{\Lambda_{1}, \ldots, \Lambda_{s}\right\}$ and $\theta$, respectively; $\Lambda_{l}$ is the least common multiple of the numbers $2 \pi / x_{i}, 2 \pi / k_{c l}(c=1, \ldots, d) ; \theta$ is the least common multiple of the numbers $2 \pi / v, 2 \pi / \omega_{c}(c=1, \ldots, d)$. Thus, in ( 1.7 ), besides the initial waves $\left(\omega_{c}, k_{c}\right), c=1, \ldots, d$, there are included the combinative resonance waves $\left(\omega_{c}, k_{c}\right), c=d+1, \ldots, q_{0}$

We obtain an equation for the function $w(x, t, x, \tau)$ by equating to zero the coefficient of $\mu$ upon suhstituting into the operator

$$
N\left(u^{(1)}\right)=u_{l}^{(1)}+\sum_{l=1}^{s} A^{(0)} u_{l}^{(1)}+B u^{(1)}-\mu f(x, t) P\left(u^{(1)}\right)
$$

in place of $u^{(1)}$ of the expression (1.7), taking into account the representation

$$
\begin{align*}
& P\left(u^{(1)}\right)=P(V+\mu w)=P(V)+\mu \varphi(x, t, a, w)  \tag{1.8}\\
& V=\sum_{e=1}^{q} a_{c}(\chi, \tau) \psi_{c} \exp \left[i\left(\omega_{c} t-k_{c} x\right)\right]+c_{c} c_{c}
\end{align*}
$$

Here $\varphi(x, t, a, w)$ is a periodic function of $x$ and $t$ (a trigonometric polynomial) and a polynomial in $a, w$ and $\mu$.

With this choice of $w$, the function $u^{(1)}$ will be an exact solution of the system

$$
\begin{align*}
& u_{t}^{(1)}+\sum_{l=1}^{s} A^{(l)} u_{l}^{(1)}+B u^{(1)}=\mu f(x, t) P\left(u^{(1)}\right)+\mu^{2} g^{(1)}(x, t, \mu)  \tag{1.9}\\
& g^{(1)}(x, t, \mu)=f(x, t) \varphi(x, t, a, w)-\frac{\partial w}{\partial \tau}-\sum_{l=1}^{s} A^{(t)} \frac{\partial w}{\partial \chi_{l}} \tag{1.10}
\end{align*}
$$

Here $f \varphi=\left\{f^{(1)} \varphi^{(1)}, \ldots, f^{(n)} \varphi^{(n)}\right\}$.
For determining $w$ we obtain the linear system

$$
\begin{align*}
& w_{i}+\sum_{l=1}^{s} A^{(l)} w_{l}+B w=h(x, t, \chi, \tau)  \tag{1.11}\\
& h(x, t, \chi, \tau)=f(x, t) P(V)-\left\{\sum_{c=1}^{q} \exp \left[i\left(\omega_{c} t-k_{c} x\right)\right] \times\right.  \tag{1.12}\\
& \left.\quad\left[I \frac{\partial a_{c}}{\partial \tau}+\sum_{i=1}^{s} A^{(l)} \frac{\partial a_{c}}{\partial \chi_{l}}\right] \psi_{c}+c_{*} c_{\bullet}\right\}
\end{align*}
$$

In the system (1.11) we regard the slow variables $\chi$ and $\tau$ as parameters, independent of $x$ and $t$. It follows from (1.12) that $h(x, t, \chi, \tau)$ consists of two terms: a product of $f$ by the trigonometric polynomial $P(V)$, in which the coefficients of the harmonics are polynomials in the complex amplitudes $a_{c}$, and a trigonometric resonace polynomial, the coefficients of the harmonics of which depend on $\partial a_{c} / \partial x$ and $\partial a_{c} / \partial \tau$. We write $h(x, t, \chi, \tau)$ in the form of a Fourier series and extract from it the resonance part

$$
\begin{gather*}
h(x, t, \chi, \tau)=\left\{\sum_{c=1}^{q} F_{c}(\chi, \tau) \exp \left[i\left(\omega_{c} t-k_{c} x\right)\right]+\right.  \tag{1.13}\\
\left.\sum_{a} F_{a}(\chi, \tau) \exp \left[i\left(\omega_{a} t-k_{a} x\right)\right]\right\}+ \text { к. с. }
\end{gather*}
$$

Here

$$
\begin{align*}
& \left(\omega_{c}, k_{\mathrm{c}}\right) \in \Omega, \quad \omega_{a}=m_{0}^{(b)} v+n^{(b)} \omega \\
& k_{\mathrm{a}}=m^{(b) x}+n^{(b) k}, \quad\left\{m^{(b)}, n^{(b)}\right\} \subset N_{1} \\
& F_{\mathrm{c}, \mathrm{a}}(x, \tau)=\frac{1}{0 \cdot \Lambda_{1} \cdots \Lambda_{s}} \int_{0}^{0} \int_{0}^{0} h(x, t, \chi, \tau) \times  \tag{1.14}\\
& \quad \exp \left[-i\left(\omega_{c, a} t-k_{c, a} x\right)\right] d t d x
\end{align*}
$$

We seek a solution of the system (1,11) in the form

$$
\begin{gather*}
w(x, t, \chi, \tau)=\left\{\sum_{c=1}^{q} w_{c}(\chi, \tau) \exp \left[i\left(\omega_{c} t-k_{\mathrm{c}} x\right)\right]+\right.  \tag{1.15}\\
\left.\sum_{a} w_{a}(\chi, \tau) \exp \left[i\left(\omega_{a} t-k_{a} x\right)\right]\right\}+
\end{gather*}
$$

Upon substituting Eqs. (1.13) and (1.15) into (1.11), we obtain linear algebraic systems for determining $w_{c, a}(\chi, \tau)$ :

$$
\begin{equation*}
H_{c, a} w_{c, a}=F_{c, a} \tag{1.16}
\end{equation*}
$$

Since $D\left(\omega_{a}, k_{a}\right) \neq 0$, we can determine $w_{a}(\chi, \tau)$ uniquely. For the existnce of $w_{c}(\chi, \tau)$ it is necessary and sufficient that

$$
\begin{equation*}
\left(\zeta_{c}, F_{c}\right)=0, \quad c=1, \ldots, q \tag{1.17}
\end{equation*}
$$

Here $\zeta_{e}$ is an arbitrary right zero vector of the matrix $H_{e}{ }^{*}$, conjugate to the matrix $H_{c}$. We restrict the discussion to the case in which the rank of the matrix $H_{c}$ is equal to $n-1$. Then $\zeta_{c}$ is determined to within a factor and the solution of the system (1.16) has the form

$$
\begin{equation*}
w_{c}(\chi, \tau)=z_{c}(\chi, \tau)+c_{c}(\chi, \tau) \psi_{c} \tag{1.18}
\end{equation*}
$$

Here $z_{c}(\chi, \tau)$ is an arbitrary solution of $(1,16)$ and $c_{c}(\chi, \tau)$ are arbitrary functions of $\chi$ and $\tau$. In what follows we assume that these functions belong to the class $C_{r+1}$.

We obtain differential equations for determining the slow amplitudes $a_{c}(\chi, \tau)$ from the condition (1.17), upon taking into account (1.14) and (1.12), in the form

$$
\begin{align*}
& \frac{\partial a_{c}}{\partial \tau}+\sum_{l=1}^{s} v_{e l} \frac{\partial a_{c}}{\partial x_{l}}=f_{c}(a), \quad v_{c l}=\frac{\left(\zeta_{c}, A^{(1)} \psi_{c}\right)}{\left(\zeta_{c}, \psi_{c}\right)}  \tag{1.19}\\
& f_{c}(a)=\frac{1}{\left(\zeta_{c}, \psi_{c}\right) \theta \cdot \Lambda_{1} \ldots \Lambda_{s}} \int_{0}^{\theta} \int_{0}^{\Lambda} f(x t) P(V) \exp \left[-i\left(\omega_{c} t-k_{c} x\right)\right] d t d x
\end{align*}
$$

Here $f_{c}(a)$ is a polynomial in the amplitudes $a_{c}$ of the same degree as $P$. The coefficients $v_{c l}$ are components of the group velocities $v_{c}$ of waves propagating with the amplitudes $a_{\mathrm{e}}$. To show this, we differentiate the identity ( $\zeta, H \psi$ ) $=0$ with respect to $k_{l}$; in doing so, we regard $\omega$ as a function of $k_{1}, \ldots, k_{s}$ determined by the dispersion equation

$$
\left(\partial \zeta / \partial k_{l}, H \psi\right)+\left(\zeta, I \partial \omega / \partial k_{l} \psi\right)-\left(\zeta, A^{(l)} \psi\right)+\left(\zeta, H \partial \psi / \partial k_{l}\right)=0
$$

Since $H \psi=0$ and $\left(\zeta, H \partial \psi / \partial k_{l}\right)=\left(H^{*} \zeta, \partial \psi / \partial k_{l}\right)=0$, then

$$
\partial \omega / \partial k_{l}=\left(\zeta, A^{(l)} \psi\right) /(\zeta, \psi)=-D_{k l^{\prime}}(\omega, k) / D_{\omega}{ }^{\prime}(\omega, k)
$$

Hence, for $k=k_{\mathrm{o}}$ we obtain $v_{\mathrm{cl}}=\partial \omega / \partial k_{l \mathrm{k}=k_{\mathrm{c}}}$. By assumption, the values of $\partial \omega / \partial k_{l \mid k=k_{c}}$ are real. From (1.5) it follows that the $v_{c l}$ are finite and the system ( 1.19 ) is hyperbolic.

The initial conditions follow from (1.6) and have the form

$$
\begin{equation*}
a_{c}(\chi, 0)=a_{c}^{(0)}, \quad c=1, \ldots, d ; a_{c}(\chi, 0)=0, \quad c=d+1, \ldots, q \tag{1.20}
\end{equation*}
$$

The system (1.19), with the initial conditions (1.20), has a locally analytic solution. We assume that the problem (1.19), (1.20) has a solution in the class $C_{\mathrm{r}+2}$ in the domain $-\infty<\chi<\infty, 0<\tau<\infty$. Then the function $w(x, t, \chi, \tau)$, defined by
the series ( 1.15 ), belong to the class $C_{r+1}$ in the ( $x, t$ )-halfspace.
The right-hand sides of the systems (1.16) are polynomials in $a, \partial a / \partial \tau$ and $\partial a /$ $\partial \chi$. and are therefore $r+1$ times continuously differentiable. Consequently, with the function $c_{\mathrm{c}}(\chi, \tau)$ chosen from the class $C_{r+1}$, the Fourier coefficients $w_{c, a}(\chi, \tau)$ also belong to the class $C_{r+1}$.

We note that the ambiguity in determing $w(x, t, \chi, \tau)$, owing to the arbitrariness in $c_{c}(\chi, \tau)$, is of the order of $\mu^{2}$ in an arbitrary bounded domain $G:\{x \in X, 0<t<T$ ( $X$ is the bounded portion of the space $\left\{x_{1}, \ldots, x_{8}\right\}$ ). In fact

$$
c_{c}(\chi, \tau)-c_{c}(\chi, 0)=\int_{0}^{\tau} \frac{\partial c_{c}}{\partial \tau} \partial \tau, \quad\left|c_{c}(\chi, \tau)-c_{c}(\chi, 0)\right| \gtrless \mu T L_{2}
$$

Here $L_{2}$ is a constant which bounds $\partial c / \partial \tau$ in the domain $G_{\mu}:\{\chi \in \mu X, 0<\tau ₹ \mu T\}$. Therefore, the ambiguity in determining $u^{(1)}$ is of the order of $\mu^{2}$. It may be shown similarly that the ambiguity in determining $\partial^{j} u^{(1)} / \partial \chi^{j}, j=1, \ldots, r$ is of the same order.

Convergence of the series (1.15) and the possibility of termwise differentiation of it $r+1$ times is determined from its nonresonance part. Differentiating systems (1.16) and solving them, we obtain

$$
D_{j} w_{a}=H_{a}^{-1} D_{j} F_{a}, \quad j=0,1, \ldots, r+1
$$

Here, the symbol $D_{j}$ stands for the derivative of order $j$ with respect to an arbitrary combination of the independent variables $\tau, \chi_{1}, \ldots, \chi_{8}$. By virtue of the condition $2^{\circ}$ there exists a constant $L_{3}$ such that

$$
\left|D_{j} w_{a}\right| \gtrless L_{3}\left|D_{j} F_{a}\right|
$$

The coefficients $F_{a}$ represent the product of the Fourier coefficients $f_{m}$ of the function $f(x, t)$ by polynomials in the amplitudes and their derivatives up to the order $r+1$ with a finite set of coefficients. In the domain $G_{\mu}$ these polynomials are uniformly bounded by a constant. Therefore, the $D_{j} w_{a}$ decrease in the same way as the Fourier corfficients of the function $D_{j} f(x, t)$, i. e. they decrease faster than

$$
\begin{aligned}
& 1 /|m|^{r+3-j}, \quad|m|^{j}=\left|m_{0}\right|^{j_{0}} \cdot\left|m_{1}\right|^{j_{1}} \ldots\left|m_{s}\right|^{j_{s}} \\
& j_{0}+j_{1}+\ldots+j_{s}=j
\end{aligned}
$$

Consequently, the series (1.15) and its derivatives to the order $r+1$ converge absolutely and uniformly in the domain $G$ and $w(x, t, \chi, \tau) \in C_{r+1}$.

It follows from (1.7) that since $a_{c} \in C_{r+2}$ and $w \in C_{r+1}$, then $w^{(1)}(x, t) \in C_{r+1}$, and from $\mathrm{Eq}_{0}(1.10)$ it follows that $g^{(1)}(x, t, \mu) \in C_{r}$.

Thus, the problem of obtaining an approximate solution is reduced to the integration of the almost linear system (1.19) and the solution of the algebraic systems (1.16). The system (1.19) is considerably simpler than the original system and its solution can be found, for example, by the method of successive approximations by integrating along the rays

$$
\begin{aligned}
& \chi_{l}-v_{c l} \tau=\text { const } \\
& a_{c}^{(m+1)}(\chi, \tau)=a_{c}^{(0)}+\int_{0}^{\tau} f_{c}\left\{a^{(m)}\left[\chi-v_{c}(\tau-\sigma), \sigma\right]\right\} d \sigma
\end{aligned}
$$

2. We show now that the approximate solution $u^{(1)}(x, t)$ is close to the exact solu-
tion $u(x, t)$ in a bounded domain of the $(x, t)$-halfspace when the initial conditions coincide or are in close proximity.

Theorem. Let $u(x, t)$ be a solution of the system (1.1), $u^{(1)}(x, t)$ be a solution of the system (1.9), and $u(x, t) \in C_{r}, u^{(1)}(x, t) \in C_{r}$ for $-\infty<x<\infty, 0<$ $t<\infty$. Let

$$
\begin{align*}
& u(x, 0)=\varphi(x), \quad u^{(1)}(x, 0)=\psi(x), \quad \varphi, \psi \in C_{r},-\infty<x<\infty  \tag{2,1}\\
& |\varphi(x)-\psi(x)|<\mu^{2} K, \quad\left|D_{k} \varphi(x)-D_{k} \psi(x)\right|<\mu^{2} K \tag{2.2}
\end{align*}
$$

Here $k \leqslant r$; $D_{k}$ is the derivative of order $k$ with respect to an arbitrary conbination of the variables $x_{1}, \ldots, x_{s} ; K$ is a constant.

Then for any $T>0$ and any bounded closed simply-connected domain $X$ with a smooth boundary $L$, there exists a constant $M$ such that

$$
\begin{aligned}
& \left|u^{(1)}(x, t)-u(x, t)\right|<\mu^{2} M, \quad \forall x \in X, 0 \gtrless t \gtrless T \\
& 0 \gtrless \mu \gtrless \mu_{0}
\end{aligned}
$$

To ease the writing we shall not indicate here the dependence of $u, u^{(1)}, g^{(1)}, \varphi$ and $\psi$ on $\mu$. We assume that the initial functions $\varphi(x, \mu)$ and $\psi(x, \mu)$, and their derivatives $D_{k} \varphi(x, \mu)$ and $D_{k} \psi(x, \mu), k \leqslant r$ are continuous functions of $\mu$. By a theorem on the continuous dependence of the solution of an almost linear system on a parameter [5], the functions $u(x, t, \mu), D_{k} u, u^{(1)}(x, t, \mu)$ and $D_{k} u^{(1)}$ are continuous functions of $\mu$ for $k \leqslant r$. The function $g^{(1)}(x, t, \mu)$, as a consequence of (1.10), also depends continuously on $\mu$.

Proof. Differentiation of the systems (1.1) and (1.9) with respect to an arbitrary combination of the variables up to the order $r$ inclusive leads to extended systems of the form

$$
\begin{align*}
& y_{t}+\sum_{l=1}^{s} R^{(l)} \frac{\partial y}{\partial x_{l}}+Q y=\mu f_{1}(x, t, y)  \tag{2.3}\\
& y_{t}^{(1)}+\sum_{l=1}^{s} R^{(l)} \frac{\partial y^{(1)}}{\partial x_{l}}+Q y^{(1)}=\mu f_{1}\left(x, t, y^{(1)}\right)+\mu^{2} g_{1}(x, t)
\end{align*}
$$

Here $y$ and $y^{(1)}$ are vectors whose components are components of the vectors $u$ and $u^{(1)}$ respectively, and their derivatives up to order $r$ inclusive, with respect to an arbitrary combination of the variables $x_{1}, \ldots, x_{s} ; R^{(b)} Q$ are cell-like square matrices formed from the matrices $A^{l}$ amd $B$, respectively. The vector-valued function $f_{1}$ is the sum of products of periodic functions of $x$ and $t$ by polynomials in $y$. The components of the vector-valued function $g_{1}(x, t)$ are components of the vector-valued function $g^{(1)}(x, t)$ and its derivatives up to order $r$ inclusive, with respect to an arbitrary combination of the variables $x_{1}, \ldots, x_{\mathrm{s}}$.

From the relations (2.1) and (2.2), we have $\left|y(x, 0)-y^{(1)}(x, 0)\right|<\mu^{2} K$. We now consider the systems (2.3) as identities in $x$ and $t$, obtained by substituting into the equations (2.3) their solutions (these solutions are assumed to satisfy the initial conditions given by the relations (2.1)). We subtract the first of the relations (2.3) from the second and consider the resulting identlty. By a lemma of Hadamard on finite increments, we have

$$
\begin{aligned}
& f_{1}\left(x, t, y^{(1)}\right)-f_{1}(x, t, y)=\varepsilon\left(x, t, y^{(1)}, y\right)\left(y^{(1)}-y\right)=\varepsilon(x, t) v(x, t) \\
& v=y^{(1)}-y
\end{aligned}
$$

The identity so obtained now assumes the form

$$
\begin{equation*}
v_{t}+\sum_{l=1}^{s} R^{(l)} \frac{\partial v}{\partial x_{l}}+(Q-\mu \mathrm{e}) v=\mu^{2} g_{1}(x, t) \tag{2.4}
\end{equation*}
$$

The elements of the matric function $\varepsilon(x, t)$ are sums of products of derivatives of the periodic function $f(x, t)$ up to order $r$ inclusive, by polynomials in $y^{(1)}$ and $y$, and they are, therefore, continuous functions in the ( $x, t$ )-halfspace, Since the function $\varepsilon(x, t)$ is assumed to be known, we can consider (2.4) as a linear system and use the results given by Courant in [5] for estimating the norm of the solution of a symmetric hyperbolic system in terms of the norm of the initial function $v(x, 0)$ and the norm of the right-hand side. We consider the conoid of dependence for the set of points of the space $\left(t, x_{1}, \ldots, x_{s}\right)$ for $0<t \geqslant T\left(x_{1}, \ldots, x_{s}\right) \in X$. Let $H(h)$ be the section of this conoid by the plane $t=h$. We then have

$$
\begin{equation*}
\|v(h)\|^{2}==\int_{R(h)} v^{2}(x, h) d x<C_{1}\|v(0)\|^{2}+C_{2} \mu^{4} \int_{0}^{h}\left\|g_{1}(\tau)\right\| d \tau \tag{2.5}
\end{equation*}
$$

By virtue of (2.2), we have $\|v(0)\|^{2}<\mu^{4} C_{3}$ therefore, from the relation (2.5) we have

$$
\|v(h)\|^{2}<\mu^{4} C_{4}, 0<h<T
$$

By Sobolev's imbedding theorem [6] it then follows that

$$
\left|u^{(1)}(x, t)-u(x, t)\right|<\mu^{2} M, 0<t<T, x \in X
$$

This completes the proof of the theorem.
The magnitude of the constant $M$ depends on the measure of the set $X$ (being proportional to it); it contains the factor $e^{\alpha T}$, where $\alpha$ depends on the coefficients of the system and characterizes the stability of the system; it also depends on the maximum of the function $g^{(1)}(x, t)$.

The author thanks G. M. Zhislin for useful remarks.

## REFERENCES

1. Bogoliubov, N. N, and Mitropil'skii, Iu. A., Asymptotic Methods in the Theory of Nonlinear Oscillations. (English translation), Gordon and Breach, New York, 1964.
2. Rabinovich, M. I., On an asymptotic method in the theory of nonlinear oscillations of distributed systems. Dokl, Akad, Nauk SSSR, Vol. 191, № 6, 1970.
3. Rabinovich, M. I. and Rozenblium, A. A., On asymptotic methods for the solution of nonlinear partial differential equations. PMM Vol. 36, $\mathrm{N}^{2} 2,1972$.
4. Mitropol'skii, Iu, A., Method of Averaging in Nonlinear Mechanics. "Naukova Dumka", Kiev, 1971.
5. Courant, R, and Friedrichs, K., Methods of Mathematical Physics, Vol. 2 : Partial Differential Equations, Wiley-Interscience, New York, 1962.
6. Sobolev, S. L. , Applications of Functional Analysis in Mathematical Physics. (English translation), American Math. Soc. Translation, Vol. 7. American Math. Soc. , Providence, 1963.
