

Orthogonally Invariant Measures and Best Approximation of Linear Operators

NIELS JUUL MUNCH*

*Matematisk Institut, Aarhus Universitet,
Ny Munkegade, DK-8000 Aarhus C, Denmark*

Communicated by T. J. Rivlin

Received July 27, 1987

This paper studies optimal information and optimal algorithms in Hilbert space for an n -dimensional average case model. The error in approximating a linear operator is the average of some error criterion E with respect to an *orthogonally invariant* measure. The orthogonally invariant measures are characterized and the problem of best approximation is solved for a wide range of error criteria E . In addition it is shown that *adaption does not help*. © 1990 Academic Press, Inc.

1. INTRODUCTION

This paper is concerned with the general problem of estimating the action of a linear bounded operator A on a real, separable Hilbert space \mathcal{H} when only finite information is available. Here *information* is provided by a map N from \mathcal{H} into the space \mathbb{R}^n of fixed finite dimension n . Knowing $Nf, f \in \mathcal{H}$, one seeks, the best recovery of Af by means of an *algorithm* φ , that is, a map $\varphi: \mathbb{R}^n \rightarrow \mathcal{H}$. In other words the difference $A - \varphi N$ should be as small as possible in a specified sense.

For a worst case error criterion this setup has been examined in [4, 5, 8–10] and others. Here we relate to an *average error* criterion as in [6–8, 12–16]. Assuming μ to be a Borel probability measure on \mathcal{H} with mean zero and finite second moment $\int_{\mathcal{H}} \|f\|^2 d\mu(f)$, the error to be minimized is

$$e(\varphi, N) = \int_{\mathcal{H}} E(Af - \varphi Nf) d\mu(f) \tag{1}$$

for some function $E: \mathcal{H} \rightarrow \mathbb{R}_+ = [0, \infty[$. The classical choice for E is $E(f) = \|f\|^2$, which constitutes the average squared error. But also the

* Research supported by the Danish Natural Science Research Council.

probabilistic or hit-and-miss criterion conforms to this framework if we choose $E(f) = 1_{[e, \infty[}(\|f\|)$.

It is the striking result of [14] (see also [6, 7, 13, 15, 16]) that for the average squared error and a certain class of "orthogonally invariant" measures μ *adaptive* linear information is not superior to *non-adaptive* linear information. For such μ , possessing a high degree of spatial symmetry, the minimal error $e(\varphi, N)$ can be obtained even within the class of non-adaptive linear information operators N and corresponding linear spline algorithms φ .

Given the implications of this result it becomes of interest to determine its precise range of validity, i.e., to determine which measures μ are orthogonally invariant. Examples from [14] are Gaussian measures and, in finite dimensions, measures μ absolutely continuous with respect to Lebesgue measure m so that the Radon-Nikodym derivative $d\mu/dm$ is rotation invariant in a suitably perturbed inner product. It turns out that these examples are quite characteristic. Denote by C the covariance operator of μ and assume that the range of C is infinite dimensional. In that case μ is orthogonally invariant if and only if it has a representation

$$\mu = \int_0^\infty \mu_{tC} dv(t), \tag{2}$$

where μ_K is the Gaussian measure on \mathcal{H} with mean zero and covariance operator K and ν is a Borel probability measure on \mathbb{R}_+ . Further the representation (2) is uniquely determined under the condition $1 = \int_0^\infty t dv(t)$ in which case we denote the measure μ by μ_C^ν . One may note from (2) that the projection of μ_C^ν onto any finite dimensional subspace of \mathcal{H} invariably is absolutely continuous w.r.t. Lebesgue measure.

The error criterion (1) has a simple probabilistic interpretation. When X is a second order random variable taking values in \mathcal{H} and μ is the induced distribution on \mathcal{H} the error $e(\varphi, N)$ is the expected value of $E((A - \varphi N)X)$. Of course Gaussian measures μ arise from Gaussian variables X . But suppose that X is in fact an instance, say X_1 , of a stochastic process $\{X_t\}_{t \geq 0}$, where each X_t has the distribution μ_{tC} . If the observation time T is subject to noise it may differ from its nominal value $T=1$, transforming the variable $X=X_1$ into $X=X_T$, where $X_T(\omega) = (X_{T(\omega)})(\omega)$ for each outcome ω . Under the hypothesis of statistical independence of $\{X_t\}_{t \geq 0}$ and T , the distribution of $X=X_T$ is μ_C^ν , where ν is the distribution of T . Thus even starting in a purely Gaussian setting it is possible to arrive naturally at measures of the type μ_C^ν .

In Section 2 of this paper we prove the structure theorem (2) and derive some ancillary properties of orthogonally invariant measures. As I have recently become aware it appears from a remark in [16, p. 363] that the

connection between orthogonally invariant measures and the “elliptically contoured” measures [3] of the form (2) has been noted previously by Kwapien in private communication. However, since the present approach is quite different from the one in [3] and the results are sharper, I feel it is justified to present this material.

Next, in Section 3, we study the approximation problem for orthogonally invariant measures. For a quite general class of error criteria E it is proved that adaptive information N is not more powerful than non-adaptive information and that for a given non-adaptive N the natural spline algorithm is optimal. In particular there is a linear optimal algorithm. The set of E 's considered includes the average squared error and the probabilistic error. Consequently our approach unifies and improves previous results for orthogonally invariant measures and the squared error [12, 14, 15] and Gaussian measures and more general criteria [13, 16, (6), (7)]. In addition a number of new results and uniqueness results are obtained.

2. ORTHOGONALLY INVARIANT MEASURES

Let \mathcal{H} be a real Hilbert space of finite or countable dimension. Consider on \mathcal{H} a Borel probability measure μ with mean zero, finite second moment $\int_{\mathcal{H}} \|f\|^2 d\mu(f)$, and covariance operator C_μ defined by

$$C_\mu = \int_{\mathcal{H}} (f \otimes f) d\mu(f).$$

Here the Hilbert-Schmidt operators on \mathcal{H} are identified with the tensor product $\mathcal{H} \otimes \mathcal{H}$ through $(f_1 \otimes f_2)g = (g, f_1)f_2$. It is assumed that C_μ is injective and that μ is symmetric, i.e., $\int_{\mathcal{H}} F(f) d\mu(f) = \int_{\mathcal{H}} F(-f) d\mu(f)$.

Following Wasilkowski and Wozniakowski [14] we define the symmetric measure μ to be *orthogonally invariant* if $\mu = \mu \circ Q_f^{-1}$ for all $f \in \mathcal{H}$ normalized so that $(C_\mu f, f) = 1$. Here Q_f is the operator $Q_f = 2(f \otimes C_\mu f) - I$ which satisfies $Q_f^2 = I$ provided $(C_\mu f, f) = 1$.

Recall that the Fourier transform or characteristic functional $\hat{\mu}$ of μ is the function from \mathcal{H} into \mathbb{C} defined by

$$\hat{\mu}(f) = \int_{\mathcal{H}} \exp(i(g, f)) d\mu(g), \quad f \in \mathcal{H},$$

and that $\hat{\mu}$ determines μ uniquely [11, pp. 11]. For any non-zero vector g in \mathcal{H} denote by l_g the functional $l_g(f) = (C_\mu g, g)^{-1/2} (f, g)$, $f \in \mathcal{H}$.

PROPOSITION 2.1. *For a symmetric measure μ the following are equivalent.*

- (a) The measure μ is orthogonally invariant.
- (b) All measures $\mu \circ l_g^{-1}$, $g \in \mathcal{H} \setminus \{0\}$, are equal.
- (c) There is a function $g: \mathbb{R}_+ \rightarrow \mathbb{R}$ so that

$$\hat{\mu}(f) = g((C_\mu f, f)^{1/2}), \quad f \in \mathcal{H}.$$

- (d) There is a twice continuously differentiable, positive definite function $g_\mu: \mathbb{R} \rightarrow \mathbb{R}$ such that

$$\hat{\mu}(f) = g_\mu((C_\mu f, f)^{1/2}), \quad f \in \mathcal{H}$$

and

$$g_\mu(0) = 1, \quad g'_\mu(0) = 0, \quad g''_\mu(0) = -1.$$

Further if μ is orthogonally invariant and if for some function $g: \mathbb{R}_+ \rightarrow \mathbb{R}$ and self-adjoint operator C it holds that $\hat{\mu}(f) = g((Cf, f)^{1/2})$, $f \in \mathcal{H}$, then there is a constant $\gamma > 0$ such that $C_\mu = \gamma^2 C$ and $g(s) = g_\mu(\gamma s)$, $s \geq 0$.

Proof. (a) \Rightarrow (c). Assume $(C_\mu f_1, f_1) = (C_\mu f_2, f_2)$. Let g be the vector $g = (C_\mu(f_1 + f_2), f_1 + f_2)^{-1/2} (f_1 + f_2)$. As $(C_\mu(f_1 + f_2), f_1 + f_2) = 2((C_\mu f_1, f_1) + (C_\mu f_1, f_2))$ one may verify that $Q_g^* f_1 = f_2$. Consequently $\hat{\mu}(f_1) = \mu \circ Q_g^{-1}(f_1) = \hat{\mu}(Q_g^* f_1) = \hat{\mu}(f_2)$.

(c) \Rightarrow (a). It is straightforward to verify the relation $Q_f C_\mu Q_f^* = C_\mu$. Thus

$$\begin{aligned} \widehat{\mu \circ Q_f^{-1}}(g) &= \hat{\mu}(Q_f^* g) \\ &= g((C_\mu Q_f^* g, Q_f^* g)^{1/2}) \\ &= g((C_\mu g, g)^{1/2}) \\ &= \hat{\mu}(g), \quad g \in \mathcal{H}. \end{aligned}$$

(c) \Leftrightarrow (b). This equivalence is seen from

$$\widehat{\mu \circ l_g^{-1}}(s) = \hat{\mu}(s(C_\mu g, g)^{-1/2} g), \quad s \in \mathbb{R}. \tag{3}$$

(c) \Rightarrow (d). Denote the common value of $\mu \circ l_g^{-1}$ by $\bar{\mu}$. It is apparent from (3) that $\bar{\mu}(s) = g(s)$, $s \geq 0$. Now (d) is simply the statement that $g_\mu = \bar{\mu}$ is the transform of a probability measure with mean zero and second moment one.

To prove the final statement of the proposition assume that

$$\hat{\mu}(f) = g_\mu((C_\mu f, f)^{1/2}) = g((Cf, f)^{1/2}), \quad f \in \mathcal{H}.$$

Consider any non-zero vector g in \mathcal{H} and put $\alpha^2 = (C_\mu g, g)$, $\beta^2 = (Cg, g)$. Then $\hat{\mu}(sg) = g_\mu(|s|\alpha) = g(|s|\beta)$. If $\beta = 0$ then $\hat{\mu}(\mathbb{R} \cdot g) = \{1\}$ and μ is con-

centrated on the orthogonal complement of g , contradicting the standing assumption that C_μ is injective. Thus $g(s) = g_\mu(\gamma s)$, $s \geq 0$, holds with $\gamma = \alpha/\beta$. Since the identity $g_\mu(\gamma \cdot) = g$ can be true for at most one value of γ it follows that

$$(C_\mu g, g) = (\gamma^2 Cg, g), \quad g \in \mathcal{H}.$$

The equality $C_\mu = \gamma^2 C$ is seen by polarization. ■

As stated in the introduction we denote by μ_C the Gaussian measure on \mathcal{H} with mean zero and covariance operator C . Similarly μ_C^ν denotes the measure given by

$$\mu_C^\nu(\mathcal{B}) = \int_0^\infty \mu_{tC}(\mathcal{B}) \, d\nu(t)$$

for all Borel sets \mathcal{B} .

THEOREM 2.2. *Let μ be an orthogonally invariant measure on an infinite dimensional, separable real Hilbert space \mathcal{H} . Then there is a Borel probability measure ν on \mathbb{R}_+ with $1 = \int_0^\infty t \, d\nu(t)$ and positive nuclear operator $C = C_\mu$ such that $\mu = \mu_C^\nu$. The pair (C, ν) is unique.*

Proof. By the proposition we can express $\hat{\mu}$ as $\hat{\mu}(f) = g_\mu((C_\mu f, f)^{1/2})$, $f \in \mathcal{H}$. Since g_μ is continuous, $\hat{\mu}$ is positive definite, and C_μ has dense range it follows that the function $g_\mu(\|f\|)$ is positive definite on \mathcal{H} . Hence by a famous theorem of Schoenberg [2, p. 152] the function $g_\mu(\sqrt{t})$ for $t \geq 0$ is the Laplace transform $\mathcal{L}\nu$ of a Borel probability measure ν on \mathbb{R}_+ . For convenience we express this as

$$g_\mu(t) = \int_0^\infty \exp(-t^2 s/2) \, d\nu(s), \quad t \geq 0. \tag{4}$$

In turn (4) implies

$$\begin{aligned} \hat{\mu}(f) &= \int_0^\infty \exp(-\frac{1}{2}(C_\mu f, f)t) \, d\nu(t) \\ &= \int_0^\infty \widehat{\mu_{tC_\mu}}(f) \, d\nu(t), \quad f \in \mathcal{H}. \end{aligned}$$

Hence $\hat{\mu}$ equals the transform of the well defined mixed measure $\mu_{C_\mu}^\nu$ and the two must be equal. Since

$$C_\mu = \int_0^\infty (tC_\mu) \, d\nu(t)$$

clearly $1 = \int_0^\infty t \, d\nu(t)$.

If $\mu = \mu_C^{\bar{v}}$ is some other representation one finds

$$\hat{\mu}(f) = g_{\mu}((C_{\mu}f, f)^{1/2}) = \bar{g}((\bar{C}f, f)^{1/2}),$$

where $\bar{g}(\sqrt{2t}) = (\mathcal{L}\bar{v})(t)$ and $g_{\mu}(\sqrt{2t}) = (\mathcal{L}v)(t)$, $t \geq 0$. The desired identification $(\bar{v}, \bar{C}) = (v, C_{\mu})$ follows from combining $\bar{g}''(0) = -\int_0^{\infty} t d\bar{v}(t) = -1$ with the proposition above and the injectivity of the Laplace transform. ■

It is apparent that the projection $\mu \circ p^{-1}$ of $\mu = \mu_C^{\bar{v}}$ onto a finite dimensional subspace of \mathcal{H} is absolutely continuous w.r.t. Lebesgue measure m with a Radon–Nikodym derivative $d(\mu \circ p^{-1})/dm$ which is \mathcal{C}^{∞} outside zero. If v vanishes in a neighbourhood of zero $d(\mu \circ p^{-1})/dm$ even belongs to the Schwartz space \mathcal{S} . In contrast, for any finite dimension, normalized integration over the boundary of the unit ball is an orthogonally invariant measure singular w.r.t. Lebesgue measure. In dimension one this is $\mu = \frac{1}{2}(\delta_1 + \delta_{-1})$ with transform $\hat{\mu}(t) = \cos(t) = \cos((t^2)^{1/2})$ which is not even positive.

The two corollaries to Theorem 2.2 and Proposition 2.1 demonstrate that the Gaussian measures have properties which are quite distinct from those of a general orthogonally invariant measure.

COROLLARY 2.3. *If an orthogonally invariant measure μ is a product measure with respect to a non-trivial orthogonal decomposition $\mathcal{H} = \mathcal{H}_1 \oplus \mathcal{H}_2$, then μ is a Gaussian measure.*

Proof. Choose non-zero vectors f_i in \mathcal{H}_i and put $\beta_{ij} = (C_{\mu}f_i, f_j)$. Denote by G_{μ} the function $G_{\mu}(t) = g_{\mu}(\sqrt{t})$, $t \geq 0$, which by l'Hospital's rule satisfies

$$\lim_{t \rightarrow 0^+} G'_{\mu}(t) = -\frac{1}{2}. \tag{5}$$

By hypothesis $\hat{\mu}(\lambda_1 f_1 + \lambda_2 f_2) = \hat{\mu}(\lambda_1 f_1) \hat{\mu}(\lambda_2 f_2)$. Consequently, as β_{12} is readily shown to be zero,

$$G_{\mu}(\lambda_1^2 \beta_{11} + \lambda_2^2 \beta_{22}) = G_{\mu}(\lambda_1^2 \beta_{11}) G_{\mu}(\lambda_2^2 \beta_{22}), \quad \lambda_i \in \mathbb{R}. \tag{6}$$

In combination with (5) the functional equation (6) implies $G_{\mu}(t) = \exp(-t/2)$, $t \geq 0$. Thus $\mu = \mu_{C_{\mu}}$. ■

For any measure λ (on \mathbb{R}_+) and positive real number α denote by λ^{α} the measure

$$\int F(x) d\lambda^{\alpha}(x) = \int F(\alpha x) d\lambda(x).$$

COROLLARY 2.4. *Let μ be the convolution measure $\mu = \mu_1 * \mu_2$, where $\mu_i = \mu_{C_i}^{\bar{v}_i}$.*

Then μ is orthogonally invariant only if either μ_i are both Gaussian or the covariances C_i are proportional.

Proof. The proof is based on (b) of Proposition 2.1. Since $C_\mu = C_1 + C_2$ is known we may set out to determine when $\mu \circ l_g^{-1}$, $g \in \mathcal{H} \setminus \{0\}$, are all equal. Notation will be as in the proof of Proposition 2.1.

Now $\mu \circ l_g^{-1} = (\mu_1 \circ l_g^{-1}) * (\mu_2 \circ l_g^{-1})$ and for

$$a = (C_1 g, g)^{1/2} ((C_1 + C_2) g, g)^{-1/2}$$

one finds that

$$\begin{aligned} \widehat{\mu \circ l_g^{-1}}(s) &= \widehat{(\mu_1 \circ l_g^{-1})}(s) \cdot \widehat{(\mu_2 \circ l_g^{-1})}(s) \\ &= \widehat{(\bar{\mu}_1)^a}(s) \cdot \widehat{(\bar{\mu}_2)^{(1-a^2)^{1/2}}}(s) \\ &= \hat{\mu}_1(as) \hat{\mu}_2((1-a^2)^{1/2} s) \end{aligned}$$

Thus the requirement is that the functions

$$g_a(s) = g_1(as) g_2((1-a^2)^{1/2} s) \quad (7)$$

should be independent of the parameter a as it ranges over the closure $I = K^-$ of the set

$$K = \{ \|C_1^{1/2} g\| \cdot \|(C_1 + C_2)^{1/2} g\|^{-1} \mid g \in \mathcal{H} \setminus \{0\} \}.$$

But I is precisely the set $\{ \|C_1^{1/2}(C_1 + C_2)^{-1/2} f\| \mid \|f\| = 1 \}^-$ which in turn is identical to the square root $W^{1/2}$ of the numerical range

$$W = \{ ((C_1 + C_2)^{-1/2} C_1 (C_1 + C_2)^{-1/2} f, f) \mid \|f\| = 1 \}^-.$$

In particular, I is an interval.

In case I , and hence W , is a singleton set we find by polarization that the C_i are proportional. Otherwise we may differentiate (7) with respect to a in the interior of I to obtain

$$\begin{aligned} (1-a^2)^{1/2} g'_1(as) g_2((1-a^2)^{1/2} s) \\ = a g_1(as) g'_2((1-a^2)^{1/2} s), \quad a \in I, s \in \mathbb{R}. \end{aligned}$$

Since g_i are everywhere positive this can be rewritten as

$$\begin{aligned} (1-a^2) \frac{d}{ds} (\ln g_1(as)) \\ = a^2 \frac{d}{ds} (\ln g_2((1-a^2)^{1/2} s)), \quad a \in I, s \in \mathbb{R}. \end{aligned}$$

and consequently for any fixed a in I

$$g_1(as)^{(1-a^2)} = g_2((1-a^2)^{1/2} s)^{a^2}, \quad s \in \mathbb{R}. \tag{8}$$

On comparison with (7), with the common value of g_a denoted by \bar{g} , this yields

$$g_1(as) = \bar{g}(s)^{a^2}, \quad a \in I, s \in \mathbb{R},$$

and it follows that the function $g_1(t)$ equals the function $\exp(t^2 \ln \bar{g}(1))$ through the interval I . Since $g_1(\sqrt{\cdot})$ is the Laplace transform of a probability measure it has an analytic continuation to the open right half plane. In turn g_1 has an analytic extension to the interior of a 45° cone symmetrically including the positive real axis. Thus from uniqueness of analytic continuation and the condition $g_1''(0) = -1$ the identity $g_1(t) = e^{-t^2/2}$, $t \in \mathbb{R}_+$, follows. Hence μ_1 , and likewise μ_2 , are Gaussian. ■

Remarks. (a) To connect Corollary 2.4 with our introductory considerations regarding random variables X_T , let Z be a sum $Z = X_T + Y_S$ of two independent variables of this kind. Corollary 2.4 states that if the covariance parameters of $(X_t)_{t \geq 0}$ and $(Y_s)_{s \geq 0}$ are not proportional then Z has an orthogonally invariant distribution only if T and S are constants.

(b) One property, however, characteristic of Gaussian measures is preserved for orthogonally invariant measures $\mu = \mu_C^\nu$. When $\{e_j\}_{j=1}^\infty$ and $\{\lambda_j\}_{j=1}^\infty$ are the eigenvectors and corresponding eigenvalues of C the limit $\lim_{n \rightarrow +\infty} (1/n) \sum_{j=1}^n (f, e_j)^2 / \lambda_j$ still exists for μ almost all f in \mathcal{H} . But it need no longer be equal to one, μ a.e. In fact ν is equal to $\mu \circ \rho^{-1}$ and μ_C , $t \in \mathbb{R}_+$, are the conditional measures for $\{\rho(f) = t\}$, where

$$\rho(f) = \lim_{n \rightarrow +\infty} \frac{1}{n} \sum_{j=1}^n \frac{(f, e_j)^2}{\lambda_j}$$

can be given any value on the set of non-convergence.

3. APPROXIMATION OF LINEAR OPERATORS

This section investigates the approximation of a linear bounded operator $A: \mathcal{H} \rightarrow \mathcal{H}$ with respect to some fixed orthogonally invariant measure $\mu = \mu_C^\nu$. First it is necessary to introduce further definitions and notations.

When the Hilbert space \mathcal{H} is identified with its own dual space of functionals, an *adaptive* linear information operator $N: \mathcal{H} \rightarrow \mathbb{R}^n$ is any map of the form $Nf = (y_i)_{i=1}^n$, where $y_1 = (f, g_1)$, $y_2 = (f, g_2(y_1))$, ...,

$y_n = (f, g_n(y_1, \dots, y_{n-1}))$, and $g_i: \mathbb{R}^{i-1} \rightarrow \mathcal{H}$ are measurable functions for $1 \leq i \leq n$. Thus the i th point of evaluation is allowed to depend (measurably) on the previous $(i-1)$ outcomes. The information operator or just information N is called *non-adaptive* if the g_i are constant functions, i.e., the points of evaluation have been chosen a priori. Given an error functional E the error of an algorithm φ is defined by

$$e(\varphi, N) = \int_{\mathcal{H}} E(Af - \varphi Nf) d\mu(f)$$

and the *radius* of an information operator N is

$$r(N) = \inf_{\varphi} e(\varphi, N).$$

Without essential loss of generality it is assumed that

$$(Cg_i(y), g_j(y)) = \delta_{ij}$$

holds for almost all y in \mathbb{R}^n . For $y = (y_i)_{i=1}^n$ in \mathbb{R}^n of course $g_i(y)$ means $g_i(y_1, \dots, y_{i-1})$. Also for y in \mathbb{R}^n we adopt the notation [13]

$$m(y) = \sum_{j=1}^n y_j Cg_j(y)$$

$$\sigma(y) = \sum_{j=1}^n g_j(y) \otimes Cg_j(y)$$

and

$$S(y) = (I - \sigma(y)) C(I - \sigma(y))^*.$$

The measure μ_C^y is transformed by N into the measure μ_f^y on \mathbb{R}^n . This is readily verified when the g_i constantly equal suitably normalized eigenvectors for C ; the general case then follows from [14, Theorem 4.2]. In [13, Theorem 3.1] it is shown that for $\mu = \mu_C$ the conditional measure for $\{Nf = y\}$ is the Gaussian measure $\mu_{m(y), S(y)}$ with mean $m(y)$ and covariance $S(y)$, i.e.,

$$\mu_C = \int_{\mathbb{R}^n} \mu_{m(y), S(y)} d\mu_f(y) \quad (9)$$

with each $\mu_{m(y), S(y)}$ supported on $\{f | Nf = y\}$. The next proposition determines the corresponding resolution of an orthogonally invariant μ with respect to N .

Denote by W_s the density function $W_s(y) = (2\pi s)^{-n/2} \exp(-\|y\|^2/2s)$, $y \in \mathbb{R}^n$, and by W^v the Radon–Nikodym derivative

$$W^v(y) = \frac{d\mu_I^v}{dm}(y) = \int_0^\infty W_s(y) dv(s).$$

PROPOSITION 3.1. *It holds that*

$$\mu_C^v = \int_{\mathbb{R}^n} \mu^y d\mu_I^v(y), \tag{10}$$

where each probability measure

$$\mu^y = W^v(y)^{-1} \int_0^\infty \mu_{m(y), sS(s^{-1/2}y)} W_s(y) dv(s) \tag{11}$$

is supported on $\{f | Nf = y\}$.

Proof. Application of (9) to the covariance operators $\bar{C} = sC$ and the informations \bar{N} given by $\bar{g}_i = s^{-1/2}g_i$ demonstrates that

$$\mu_{sC} = \int_{\mathbb{R}^n} \mu_{m(s^{1/2}y), sS(y)} d\mu_I(y)$$

and that each $\mu_{m(s^{1/2}y), sS(y)}$ is supported on $\{f | Nf = s^{1/2}y\}$. Thus

$$\mu_C^v = \int_0^\infty \int_{\mathbb{R}^n} \mu_{m(s^{1/2}y), sS(y)} d\mu_I(y) dv(s)$$

which after reshuffling, using

$$\begin{aligned} & \int_0^\infty \int_{\mathbb{R}^n} F(s, y) d\mu_I(y) dv(s) \\ &= \int_{\mathbb{R}^n} \int_0^\infty F(s, s^{-1/2}y) W_s(y) dv(s) dy, \end{aligned}$$

becomes (10) and (11). ■

In the sequel the following very general class of error functionals is considered. A measurable function $E: \mathcal{H} \rightarrow \mathbb{R}_+$ is called an *allowable* error functional if each set

$$\mathcal{B}_t = \{f \in \mathcal{H} | E(f) < t\}$$

is convex and balanced. This includes the average squared error and the error in probability. Moreover every convex function $E: \mathcal{H} \rightarrow \mathbb{R}_+$ with

$E(0)=0$ is allowable. We shall refer to E as a *standard error function* if \mathcal{B}_t has the form $\mathcal{B}_t = F(t)\mathcal{B}$, where \mathcal{B} is a bounded, convex, open set containing zero and F is a continuous bijection of \mathbb{R}_+ . In this case E is given by $E(f) = G(p_{\mathcal{B}}(f))$ for $G = F^{-1}$ and the continuous Minkowski seminorm $p_{\mathcal{B}}(f) = \inf\{t > 0 \mid f \in t\mathcal{B}\}$. The set of standard functionals includes in particular functions of the type $E(f) = H(\|f\|)$.

LEMMA 3.2. *Let E be an allowable error functional and let μ_C be a Gaussian measure. Then the function*

$$\chi(g) = \int_{\mathcal{H}} E(f - g) d\mu_C(f)$$

of g in \mathcal{H} attains its minimum value at $g=0$. If $\chi(0)$ is finite and E is a standard functional this minimum is unique.

Proof. The main tool here is the identity

$$\chi(g) = \int_{\mathcal{H}} E(f) d\mu_{g,C}(f) = \int_0^{\infty} t d(\mu_{g,C}(\mathcal{B}_t)). \quad (12)$$

Optimality of $g=0$ follows from $\mu_{g,C}(\mathcal{B}_t) \leq \mu_C(\mathcal{B}_t)$, $t \in \mathbb{R}_+$, which holds by the hypothesis on \mathcal{B}_t , cf. [13, Lemma 3.1] and [1, Theorem 1].

If conversely, $\chi(0) = \chi(g) < +\infty$ then necessarily $\mu_{g,C}(\mathcal{B}_t) = \mu_C(\mathcal{B}_t)$, $t \in \mathbb{R}_+$. For the standard case $\{\mathcal{B}_t\}_{t \geq 0}$ is equal to $\{t\mathcal{B}\}_{t \geq 0}$. From (12) and the symmetry of μ_C

$$\begin{aligned} \int_{\mathcal{H}} p_{\mathcal{B}}(f) d\mu_C(f) &= \int_{\mathcal{H}} p_{\mathcal{B}}(f - g) d\mu_C(f) \\ &= \int_{\mathcal{H}} \frac{1}{2} (p_{\mathcal{B}}(f - g) + p_{\mathcal{B}}(f + g)) d\mu_C(f). \end{aligned}$$

Combined with the convexity of $p_{\mathcal{B}}$ this implies

$$2p_{\mathcal{B}}(f) = p_{\mathcal{B}}(f - g) + p_{\mathcal{B}}(f + g), \quad \mu_C \text{ a.e.} \quad (13)$$

Take a sequence $f_n \rightarrow 0$ for which (13) holds. Then by the continuity of $p_{\mathcal{B}}$ (\mathcal{B} open), $p_{\mathcal{B}}(g) = 0$, and by the faithfulness of $p_{\mathcal{B}}$ (\mathcal{B} bounded) $g = 0$. ■

When N is non-adaptive the constant values of $S(y)$ and $g_i(y)$ are simply denoted S and g_i .

THEOREM 3.3. *Assume that N is non-adaptive information, $\mu = \mu_C^y$ is an*

orthogonally invariant measure on \mathcal{H} , and E is an allowable error functional. Then

(a) The spline algorithm

$$\varphi^s: (y_i)_{i=1}^n \rightarrow \sum_{i=1}^n y_i ACg_i \tag{14}$$

is an optimal algorithm. When $e(\varphi^s, N)$ is finite and E is a standard functional φ^s is a unique optimal algorithm.

(b) $r(N) = \int_{\mathcal{H}} E(f) d\mu_{ASA}^y(f)$.

(c) When E is p -homogeneous, i.e., $E(\alpha f) = |\alpha|^p E(f)$,

$$r(N) = \left(\int_0^\infty s^{p/2} dv(s) \right) \int_{\mathcal{H}} E(f) d\mu_{ASA}^y(f).$$

Proof. (a) Due to Proposition 3.1

$$e(\varphi, N) = \int_{\mathbb{R}^n} \int_{\mathcal{H}} E(Af - \varphi(y)) d\mu^y(f) d\mu_I^y(y), \tag{15}$$

where

$$\begin{aligned} \int_{\mathcal{H}} E(Af - \varphi(y)) d\mu^y(f) &= W^v(y)^{-1} \int_0^\infty W_s(y) \\ &\quad \times \int_{\mathcal{H}} E(Af - \varphi(y)) d\mu_{m(y),ss}(f) dv(s) \\ &= W^v(y)^{-1} \int_0^\infty W_s(y) \\ &\quad \times \int_{\mathcal{H}} E(f - (\varphi(y) - Am(y))) \\ &\quad \times d\mu_{sASA}^*(f) dv(s). \end{aligned} \tag{16}$$

From this and Lemma 3.2 it is clear that the algorithm $\varphi^s(y) = Am(y)$ (for almost all y) has the desired properties.

(b) Just combine (15), (16), and (a).

(c) This is a consequence of the general relation

$$\int_{\mathcal{H}} F(f) d\mu_{|\gamma|^2K}(f) = \int_{\mathcal{H}} F(\gamma f) d\mu_K(f).$$

It is seen from 3.3(c) that for $E(f) = p_{\mathcal{B}}(f)^p$ and other p -homogeneous functions the approximation problem for μ_C^v is equivalent to the one for μ_C .

Next we want to consider a restricted class of sets \mathcal{B} . But before we do so it is appropriate for us to touch on the problem of optimal information. Denote by R_n the operator

$$R_n = ASA^* = A(I - \sigma) C(I - \sigma)^* A^*$$

and define the n th radius of the approximation problem to be

$$r^n = \inf_N r(N).$$

In the next proposition it is tacitly assumed that all eigenvalues of ACA^* are non-degenerate. The general case is similar but more complicated to state.

PROPOSITION 3.4. *Assume that E is a standard error functional of the form $E(f) = H(\|f\|)$.*

Then $r^n = r(\bar{N})$, where the information \bar{N} is given via the n principal eigenvalues and eigenvectors (λ_i, f_i) of ACA^ through $\bar{g}_i = \lambda_i^{-1/2} A^* f_i$. If N is any information then $r(N) = r(\bar{N})$ if and only if*

$$Rg_1 + \cdots + \mathbb{R}g_n = \mathbb{R}\bar{g}_1 + \cdots + \mathbb{R}\bar{g}_n.$$

Proof. By 3.3(b) the value of $r(N)$ increases when the eigenvalues of ASA^* increase. Compute

$$\begin{aligned} R_n &= A \left(I - \sum_{i=1}^n g_i \otimes Cg_i \right) C \left(I - \sum_{i=1}^n Cg_i \otimes g_i \right) A^* \\ &= AC \left(I - \sum_{i=1}^n Cg_i \otimes g_i \right)^2 A^* \\ &= AC \left(I - \sum_{i=1}^n Cg_i \otimes g_i \right) A^* \\ &= AC^{1/2} \left(I - \sum_{i=1}^n C^{1/2}g_i \otimes C^{1/2}g_i \right) C^{1/2}A^*. \end{aligned}$$

Then R_n is given by

$$R_n = AC^{1/2}(I - P)C^{1/2}A^*,$$

where P is the orthogonal projection onto the linear span of $\{C^{1/2}g_i\}_{i=1}^n$.

The non-zero eigenvalues of $R_n = (AC^{1/2}(I-P))(AC^{1/2}(I-P))^*$ equal those of

$$\begin{aligned} \tilde{R}_n &= (AC^{1/2}(I-P))^* (AC^{1/2}(I-P)) \\ &= (I-P) C^{1/2} A^* A C^{1/2} (I-P). \end{aligned}$$

Repeating the argument we note that the (non-zero) eigenvalues of $C^{1/2} A^* A C^{1/2}$ are $\{\lambda_i\}_{i=1}^\infty$. Consequently a minimal set of eigenvalues for R_n , namely $\{\lambda_i\}_{i=n+1}^\infty$, exists and is obtained if and only if

$$\mathbb{R} C^{1/2} g_1 + \dots + \mathbb{R} C^{1/2} g_n = \mathbb{R} \eta_1 + \dots + \mathbb{R} \eta_n, \tag{17}$$

where η_i are the n principal eigenvectors of $C^{1/2} A^* A C^{1/2}$. However, η_i are proportional to $C^{1/2} A^* f_i$ and (17) is equivalent to

$$\mathbb{R} g_1 + \dots + \mathbb{R} g_n = \mathbb{R} A^* f_1 + \dots + \mathbb{R} A^* f_n. \blacksquare$$

The above proposition, which improves [13, pp. 738–741], is included at this point mainly to emphasize that the directions in \mathcal{H} determined by the eigenvectors $\{f_i\}_{i=1}^\infty$ of ACA^* have a special significance. Thus prepared the reader will hopefully admit to the relevance of the sets \mathcal{B} in the following corollaries to Theorem 3.3.

COROLLARY 3.5. *Let E be the functional $E(f) = G(p_{\mathcal{B}}(f))$, where \mathcal{B} is defined by*

$$\mathcal{B} = \left\{ f \in \mathcal{H} \mid \sum_{i=1}^\infty a_i (f, f_i)^2 < 1 \right\}$$

for some bounded set $\{a_i\}_{i=1}^\infty$ of positive numbers and G is a continuously differentiable bijection of \mathbb{R}_+ . Let θ be the real part of the function

$$\varphi(\lambda) = \prod_{j=n+1}^\infty (1 - 2i\lambda\lambda_j a_j)^{-1/2}.$$

Then

$$r(\bar{N}) = \sqrt{\frac{2}{\pi}} \int_0^\infty \int_0^\infty G((ts)^{1/2}) \hat{\theta}(t) dt dv(s), \tag{18}$$

where $\hat{\theta}$ denotes the Fourier transform.

Proof. First we claim that

$$\mu_{sA^*A}(t\mathcal{B}) = \frac{1}{\pi} \int_{-\infty}^\infty \frac{\sin(t^2\lambda)}{\lambda} \theta(s\lambda) d\lambda.$$

We shan't go into the details of this. The proof is an application of the characteristic function trick that can be found for instance in [17, pp. 66]. Now for $F = G^{-1}$

$$\begin{aligned} & \frac{d}{dt} (\mu_{sAS_A^*}(F(t)\mathcal{B})) \\ &= \frac{1}{\pi} \int_{-\infty}^{\infty} 2F'(t) F(t) \cos(F(t)^2 \lambda) \theta(s\lambda) d\lambda \\ &= \sqrt{\frac{2}{\pi}} 2s^{-1} F'(t) F(t) \hat{\theta}(s^{-1} F(t)^2) \\ &= \sqrt{\frac{2}{\pi}} \frac{d}{dt} (s^{-1} F(t)^2) \hat{\theta}(s^{-1} F(t)^2) \end{aligned}$$

and by Theorem 3.3(b)

$$\begin{aligned} r(\bar{N}) &= \sqrt{\frac{2}{\pi}} \int_0^{\infty} \int_0^{\infty} t \frac{d}{dt} (s^{-1} F(t)^2) \\ &\quad \times \hat{\theta}(s^{-1} F(t)^2) dt dv(s) \\ &= \sqrt{\frac{2}{\pi}} \int_0^{\infty} \int_0^{\infty} G((ts)^{1/2}) \hat{\theta}(t) dt dv(s). \end{aligned}$$

This is (18). ■

COROLLARY 3.6. For each $E(f) = p_{\mathcal{B}}(f)^{2p}$ it holds that

$$r(\bar{N}) = r(\bar{N}, p_{\mathcal{B}}^{2p}) = (-i)^p \varphi^{(p)}(0) \int_0^{\infty} s^p dv(s). \quad (19)$$

In particular

$$r(\bar{N}, p_{\mathcal{B}}^2) = \sum_{j=n+1}^{\infty} \lambda_j a_j$$

and

$$r(\bar{N}, p_{\mathcal{B}}^4) = \left(2 \sum_{j=n+1}^{\infty} (\lambda_j a_j)^2 + \left(\sum_{j=n+1}^{\infty} \lambda_j a_j \right)^2 \right) \int_0^{\infty} s^2 dv(s).$$

Proof. From (18)

$$r(\bar{N}, p_{\mathcal{B}}^{2p}) = \sqrt{\frac{2}{\pi}} \int_0^{\infty} t^p \hat{\theta}(t) dt \int_0^{\infty} s^p dv(s).$$

Here the second factor may or may not be finite. Our objective is to deter-

mine the value of the first factor. Since $\theta(t) = \frac{1}{2}(\varphi(t) + \varphi(-t))$ it follows that $\hat{\theta}(t) = \frac{1}{2}(\hat{\varphi}(t) + \hat{\varphi}(-t))$. Hence

$$\begin{aligned} & \sqrt{\frac{2}{\pi}} \int_0^\infty t^p \hat{\theta}(t) dt \\ &= \frac{1}{\sqrt{2\pi}} \int_0^\infty t^p (\hat{\varphi}(t) + \hat{\varphi}(-t)) dt \\ &= \frac{1}{\sqrt{2\pi}} \int_{-\infty}^\infty |t|^p \hat{\varphi}(t) dt. \end{aligned} \tag{20}$$

When p is even this is

$$\frac{(-i)^p}{\sqrt{2\pi}} \int_{-\infty}^\infty (\varphi^p)^\wedge(t) dt = (-i)^p \varphi^{(p)}(0)$$

and we are done. For odd p (20) can be rewritten as

$$\frac{(-i)^p}{\sqrt{2\pi}} \int_{-\infty}^\infty \operatorname{sgn}(t) (\varphi^p)^\wedge(t) dt. \tag{21}$$

The function or tempered distribution sgn has Fourier transform $\widehat{\operatorname{sgn}} = (-i)\sqrt{2/\pi} \operatorname{Vp}(1/t)$, where Vp denotes the Cauchy principal value. Thus (20) is equal to

$$\frac{(-i)^{p+1}}{\pi} \lim_{\epsilon \rightarrow 0+} \int_{|t| \geq \epsilon} \frac{\varphi^{(p)}(t)}{t} dt. \tag{22}$$

To estimate this integral we exploit the fact that $z^{-1/2}$ is an analytic function of z in the half plane $\{z \in \mathbb{C} \mid \operatorname{Re} z > 0\}$. Indeed $z^{-1/2} = (1/\sqrt{\pi}) \int_{-\infty}^\infty e^{-zx^2} dx$. In turn φ is analytic in the region $\{z \in \mathbb{C} \mid \operatorname{Im} z > -\gamma\}$, where $\gamma = (\max_j 2\lambda_j a_j)^{-1}$ and the integral of $\varphi(z)/z$ along the contour indicated in Fig. 1 is zero for any values of ϵ and R .

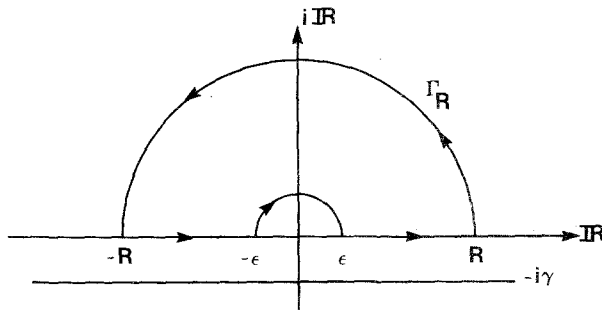


FIGURE 1

Since the integral along the semicircle Γ_R converges to zero it follows that the limit in (22) is in fact

$$\begin{aligned} & \frac{(-i)^{p+1}}{\pi} \lim_{\varepsilon \rightarrow 0+} \int_0^\pi \frac{\varphi^{(p)}(\varepsilon e^{i\theta})}{\varepsilon e^{i\theta}} i \varepsilon e^{i\theta} d\theta \\ & = (-i)^p \varphi^{(p)}(0). \end{aligned}$$

Finally, this expression is calculated for the specific values $p=1$ and $p=2$ by use of

$$\varphi'(z) = A(z) \varphi(z), \quad (23)$$

where

$$A(z) = i \sum_{j=n+1}^{\infty} \frac{\lambda_j a_j}{(1 - 2iz\lambda_j a_j)}. \quad \blacksquare$$

Remark. By iterating (23) and using

$$A^{(k)}(0) = (i)^{k+1} 2^k \cdot k! \sum_{j=n+1}^{\infty} (\lambda_j a_j)^{k+1}$$

one may of course generate any desired instance of $(-i)^p \varphi^{(p)}(0)$. But we have not been able to find a closed expression for this.

COROLLARY 3.7. For $E(f) = \|f\|^{2p}$ the n th radius $r^n = r^n(\|\cdot\|^{2p})$ of the approximation problem is

$$r^n(\|\cdot\|^{2p}) = (-i)^p \varphi^{(p)}(0) \int_0^\infty s^p dv(s),$$

where $\varphi(z) = \prod_{j=n+1}^{\infty} (1 - 2iz\lambda_j)^{-1/2}$. In particular

$$r^n(\|\cdot\|^2) = \sum_{j=n+1}^{\infty} \lambda_j$$

and

$$r^n(\|\cdot\|^4) = \left(2 \sum_{j=n+1}^{\infty} \lambda_j^2 + \left(\bar{2} \sum_{j=n+1}^{\infty} \lambda_j \right)^2 \right) \int_0^\infty s^2 dv(s).$$

Proof. Combine Corollary 3.6 and Proposition 3.4. \blacksquare

For $E(f) = \|f\|^p$ and other standard functionals of the form $E(f) = H(\|f\|)$ we may derive a rather nice expression for the optimal approxima-

tion φN . From Theorem 3.3(a) and Proposition 3.4 it follows that φN is optimal if and only if $\mathbb{R}g_1 + \dots + \mathbb{R}g_n = \mathbb{R}\bar{g}_1 + \dots + \mathbb{R}\bar{g}_n$ and φ has the form (14). Again let P denote the projection onto the linear span of $\{\bar{g}_i\}_{i=1}^n$. Since $\{C^{1/2}g_i\}_{i=1}^n$ is an orthonormal basis in $C^{1/2}P$ one finds, for any f in the domain of $C^{-1/2}$,

$$\begin{aligned} \varphi^s N(f) &= \sum_{i=1}^n (f, g_i) ACg_i \\ &= AC^{1/2} \sum_{i=1}^n (C^{-1/2}f, C^{1/2}g_i) C^{1/2}g_i \\ &= AC^{1/2}QC^{-1/2}f, \end{aligned}$$

where Q is the projection onto $C^{1/2}P = \text{span}\{\eta_i\}_{i=1}^n$ (cf. the proof of Proposition 3.4). Here the equation $\varphi^s N(f) = AC^{1/2}QC^{-1/2}f$ is independent of the choice of $\{g_i\}_{i=1}^n$. Consequently $AC^{1/2}QC^{-1/2}$ extends to a bounded operator in \mathcal{H} and this operator is the unique optimal value of φN . Finally, using $g_i = \bar{g}_i = \lambda_i^{-1/2}A^*f_i$ and the very definition of f_i , one finds

$$\begin{aligned} (\varphi N)^{\text{optimal}}(f) &= \sum_{i=1}^n (f, \lambda_i^{-1/2}A^*f_i) AC(\lambda_i^{-1/2}A^*f_i) \\ &= \left(\sum_{i=1}^n A^*f_i \otimes f_i \right) (f) \\ &= \left(\sum_{i=1}^n f_i \otimes f_i \right) Af. \end{aligned}$$

Thus $(\varphi N)^{\text{opt}}$ is the composition of A and an orthogonal projection of rank n .

Finally, in closing this paper, we turn to the problem of adaptive versus non-adaptive information. When N is (adaptive) information let $N_y, y \in \mathbb{R}^n$, be the non-adaptive information given by $g_i = g_i(y)$.

The heart of the very elegant proof in [13] that “adaption doesn’t help” is the equality $\mu^y(N) = \mu^y(N_y)$ between conditional measures. It is apparent from (11) that this does not hold generally for non-Gaussian measures μ_C^y . Nevertheless we have the following.

THEOREM 3.8. *For any allowable error functional E and any information N*

$$r(N) \geq \inf_{y \in \mathbb{R}^n} r(N_y). \tag{24}$$

Further if E is a standard error functional of the form $E(f) = H(\|f\|)$, $r(N) = r^n$ if and only if

$$\mathbb{R}g_1(y) + \dots + \mathbb{R}g_n(y) = \mathbb{R}\bar{g}_1 + \dots + \mathbb{R}\bar{g}_n \tag{25}$$

holds for almost all y in \mathbb{R}^n .

Proof. Using the results of Proposition 3.1, Lemma 3.2, and Theorem 3.3 compute

$$\begin{aligned} e(\varphi, N) &= \int_{\mathcal{H}} E(Af - \varphi Nf) d\mu(f) \\ &= \int_{\mathbb{R}^n} E(Af - \varphi(y)) d\mu^y(f) d\mu_1^y(y) \\ &= \int_{\mathbb{R}^n} \int_0^\infty W_s(y) \int_{\mathcal{H}} E(Af - \varphi(y)) \\ &\quad \times d\mu_{m(y), sS(s^{-1/2}y)}(f) dv(s) dy \\ &\geq \int_{\mathbb{R}^n} \int_0^\infty W_s(y) \int_{\mathcal{H}} E(Af) \\ &\quad \times d\mu_{sS(s^{-1/2}y)}(f) dv(s) dy \\ &= \int_0^\infty \int_{\mathbb{R}^n} \int_{\mathcal{H}} E(Af) d\mu_{sS(y)}(f) d\mu_I(y) dv(s) \\ &\geq \inf_{y \in \mathbb{R}^n} \int_0^\infty \int_{\mathcal{H}} E(Af) d\mu_{sS(y)}(f) dv(s) \\ &= \inf_{y \in \mathbb{R}^n} r(N_y). \end{aligned}$$

This proves (24).

For the final case to be considered it can be read of the above string of calculations that $r(N) = r^n$ if and only if

$$\varphi(y) = Am(y) \tag{26a}$$

$$r^n = r(N_y) \tag{26b}$$

holds for almost all y in \mathbb{R}^n . Combining (26b) with Proposition 3.4 one gets (25). The optimal algorithm is given by

$$\varphi((y_i)_{i=1}^n) = \sum_{i=1}^n y_i ACg_i(y_1, \dots, y_{i-1})$$

for almost all $y = (y_i)_{i=1}^n$ in \mathbb{R}^n . ■

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