

- [7] S. A. KOVYAZIN, *A variant of the law of large numbers for random subsets of a space with an atomless measure*, Preprint VTs SO Akad. Nauk SSSR, Krasnoyarsk, 1983, no. 1, pp. 8-9. (In Russian.)
- [8] A. I. ORLOV, *Random sets: laws of large numbers and the testing of statistical hypotheses*, Theory Prob. Appl., 23 (1978), pp.
- [9] Z. ARTSTEIN AND R. A. VITALE, *A strong law of large numbers for random compact sets*, Ann. Probab., 3 (1975), pp. 879-882.

## A GEOMETRIC APPROACH TO MAXIMUM LIKELIHOOD ESTIMATION FOR INFINITE-DIMENSIONAL GAUSSIAN LOCATION. II.

B. S. TSIREL'SON

(Translated by K. Durr)

Part I of this paper [1] examined necessary and sufficient conditions for the existence, uniqueness and consistency of the MLE for an infinite-dimensional location parameter of a Gaussian measure. Regarding the infinite-dimensional set  $V$  to which the parameter to be estimated was known to belong, nothing was assumed beforehand except closedness. The geometric tool of the study was the mean width  $h_1(V)$  introduced by Sudakov. This second part considers the more special question of the stability of the MLE with respect to sampling fluctuations. The set  $V$  is assumed to be convex. The geometric tool will be the  $k$ -th thickness  $h_k(V)$  introduced by Simone Chevet (however, the statements of the main probabilistic results contain only  $h_1$ ). Roughly speaking, it is found that in the MLE almost all degrees of freedom (except a finite number) inherent in a sample point are "frozen". This fact which is fairly natural in itself, occasionally produces unexpected and even curious effects in applications. For instance, in some cases of estimating a signal in additive white noise, the MLE turns out to be a step function regardless of the properties of the signal.

We continue to use the notation and things introduced in [1]:  $E, \gamma, E_0, \langle \theta, x \rangle, \|\theta\|, \gamma_\theta, \gamma_{\theta, \sigma}, \mathcal{L}_\sigma(\theta, x), V$ , as well as  $B(\theta, r)$ . We shall also consider the finite-dimensional case. If  $E$  is finite-dimensional, then without loss of generality we can assume that  $E = E_0 = \mathbb{R}^n$ ;  $\langle \theta, \eta \rangle$  and  $\|\theta\|$  are the usual Euclidean scalar product and norm in  $\mathbb{R}^n$ ;  $\gamma = \gamma^n$  is the standard Gaussian measure in  $\mathbb{R}^n$  with density

$$(2\pi)^{-n/2} \exp(-\frac{1}{2}\|x\|^2)$$

with respect to Lebesgue measure. The logarithm likelihood can be written as

$$\log \mathcal{L}'_\sigma(\theta, x) = (\|x\|^2 - \|x - \theta\|^2) / 2\sigma^2,$$

from which it is evident that the MLE is simply the closest point in  $V$  to  $x$  (however, in the infinite-dimensional case the vectors  $x$  and  $x - \theta$  are "infinitely long" in the considered norm).

In Theorems 1-2 below,  $V \subset \mathbb{R}^n$  is a convex closed set containing the point  $\theta$  with coordinates  $\theta_1, \dots, \theta_n$ ,  $\sigma$  is a positive number and  $\hat{\theta}(x)$  denotes the closest point in  $V$  to  $x$ .

**Theorem 1.** *Let  $V$  be a convex polyhedron, i.e., the intersection of a finite number of halfspaces, and  $F_k$  be the set of all points lying on a  $k$ -dimensional face of the polyhedron  $V$  but not lying on a face of lower dimension (here  $F_0$  is the set of vertices and  $F_n$  the interior of  $V$ ). Let  $\xi$  be a Gaussian random vector in  $\mathbb{R}^n$  whose coordinates  $\xi_k$  are independent random variables with the means  $\theta_k$  and the same variance  $\sigma^2$ .*

(a) If  $V$  is bounded, then

$$P\{\hat{\theta}(\xi) \in F_k\} \cong \frac{1}{k!} \left(\frac{C}{\sigma}\right)^k$$

for all  $k=0, 1, \dots, n$ , where  $C = (2\pi)^{-1/2} h_1(V)$ ;

(b)

$$P\left\{\hat{\theta}(\xi) \in \bigcup_{i=k}^n F_i\right\} \cong 10 \left(\frac{e^2 C_2^2}{k} \log \frac{k}{C_2^2}\right)^{k/2}$$

for  $k > C_2^2$  and  $\sigma^2 k \log(k/C_2^2) \geq 1$ ; here  $C_2 = (2\pi)^{-1/2} h_1(V \cap B(\theta, 1))$ .

**Theorem 2.** Suppose that  $\xi_1 = (\xi_{11}, \dots, \xi_{1n})$  and  $\xi_2 = (\xi_{21}, \dots, \xi_{2n})$  are Gaussian random vectors in  $R^n$  and for each  $k$  the variables  $\xi_{1k}$  and  $\xi_{2k}$  both have the mean  $\theta_k$  and the variance  $\sigma^2$ , and their correlation coefficient is equal to  $\rho$ ; and suppose that all the other correlations vanish.

(a) If  $V$  is bounded,  $C = (2\pi)^{-1/2} h_1(V)$  and  $\rho \geq 1 - \min(\sigma/5C, \frac{1}{2})$ , then for any  $u \geq 0$

$$P\left\{\frac{\|\hat{\theta}(\xi_1) - \hat{\theta}(\xi_2)\|}{\sqrt{\sigma(1-\rho)}} \geq 3\sqrt{u\sigma + 2C}\right\} \cong e^{-u};$$

(b) if  $C_2 = (2\pi)^{-1/2} h_1(V \cap B(\theta, 1))$  and

$$\rho \geq 1 - \min((5C_2(\sqrt{2(u+3)} + C_2))^{-1}, \sigma/5C_2, \frac{1}{2}),$$

then for any  $u \geq 0$

$$P\{\|\hat{\theta}(\xi_1) - \hat{\theta}(\xi_2)\|/\sqrt{\sigma(1-\rho)} \geq q\} \cong e^{-u},$$

where

$$q = 3 \max(\sqrt{\sigma}(\sqrt{u+3} + \sqrt{2} C_2), (\sigma(u+3) + 2C_2)^{1/2}).$$

The inequalities in Theorems 1-2 do not involve the dimension  $n$  of the space  $E$ ; they can of course be carried over to the infinite-dimensional case. To avoid talking about faces of a polyhedron, we introduce the following definition. For every  $\theta \in V$  we define  $K(\theta)$  to be the largest  $k$  for which there are linearly independent vectors  $\eta_1, \dots, \eta_k \in E_\theta$  such that  $\theta + a_1 \eta_1 + \dots + a_k \eta_k \in V$  for any  $a_1, \dots, a_k \in [-1, +1]$ . If  $\theta$  is an extreme point of  $V$ , then  $K(\theta) = 0$ . If any amount of such  $\eta_i$  exist, then  $K(\theta) = +\infty$ . Under the conditions of Theorem 1, obviously  $K(\theta) = k$  for  $\theta \in F_k$ . We introduce yet another definition to avoid the correlation coefficients of the components. Let  $\xi_1, \xi_2$  be random elements of the space  $E$  each with the distribution  $\gamma_{\theta, \sigma}$  and let  $\rho \in (-1, +1)$ . We say that  $\xi_1$  and  $\xi_2$  are  $\rho$ -correlated if they are representable as

$$\begin{aligned} \xi_1 &= \theta + \sigma(((1+\rho)/2)^{1/2} \zeta_1 + ((1-\rho)/2)^{1/2} \zeta_2), \\ \xi_2 &= \theta + \sigma(((1+\rho)/2)^{1/2} \zeta_1 + ((1-\rho)/2)^{1/2} \zeta_2), \end{aligned}$$

where  $\zeta_1, \zeta_2$  are independent random elements of  $E$  each with the distribution  $\gamma$ . Under the conditions of Theorem 2 the random vectors  $\xi_1$  and  $\xi_2$  are obviously  $\rho$ -correlated.

In Theorems 3-4 below,  $E$  can be both finite-dimensional and infinite-dimensional;  $V \subset E_\theta$  is a convex closed set. According to [1] the MLE of the parameter  $\theta$  ranging over  $V$  is well-defined if and only if the characteristic  $C_1(V)$  introduced in [1] is finite. In contrast to [1],  $V$  is assumed here to be convex, and in this case  $C_1(V)$  is either infinite or 0; and if it is 0, then the other characteristic  $C_2(V, \theta)$  for  $\theta \in V$  introduced in [1] becomes

$$C_2(V, \theta) = (2\pi)^{-1/2} h_1(V \cap B(\theta, 1)).$$

Indeed, for convex  $V$  we have

$$(V \cap B(\theta, r)) - \theta \subset r((V \cap B(\theta, 1)) - \theta),$$

and so also

$$(1) \quad h_1(V \cap B(\theta, r)) \leq r h_1(V \cap B(\theta, 1))$$

for  $r \geq 1, \theta \in V$ .

It is assumed in Theorems 3-4 that  $C_1(V) < +\infty$ ;  $\theta$  is a point in  $V$ ,  $\sigma$  is a positive number; and  $\hat{\theta}(x)$  is defined as in [1].

**Theorem 3.** Let  $\xi$  be a random element in  $E$  having the distribution  $\gamma_{\theta, \sigma}$ .

(a) If  $V$  is bounded, then for any  $a \geq 0$

$$E \exp(aK(\hat{\theta}(\xi))) \leq \exp(Ce^a/\sigma),$$

where  $C = (2\pi)^{-1/2}h_1(V)$ ;

(b)

$$P\{K(\hat{\theta}(\xi)) \geq k\} \leq 10(e^2 C_2^2 k^{-1} \log(k/C_2^2))^{k/2}$$

for  $k > C_2^2$  and  $\sigma^2 k \log(k/C_2^2) \geq 1$ ; here  $C_2 = (2\pi)^{-1/2}h_1(V \cap B(\theta, 1))$ .

**Theorem 4.** Suppose that  $\xi_1, \xi_2$  are random elements in  $E$  both having the distribution  $\gamma_{\theta, \sigma}$  and that they are  $\rho$ -correlated. Then parts (a) and (b) of Theorem 2 hold.

**REMARK 1.** The constant "3" in front of the radical in the inequality in part (a) of Theorem 2 can be improved. But this is all that can be done to strengthen this inequality if  $\rho$  is close to 1 and  $C/\sigma$  and  $u$  are not close to 0. More precisely, let the function  $B$  of four arguments be such that in the conditions of part (a) of Theorem 2 (or Theorem 4—it makes no difference)

$$\lim_{\rho \rightarrow 1^-} P\left\{ \frac{\|\hat{\theta}(\xi_1) - \hat{\theta}(\xi_2)\|}{\sqrt{\sigma(1-\rho)}} \geq B(C, u, \sigma, \rho) \right\} \leq e^{-u}.$$

Then it can be shown that the expression

$$\lim_{\rho \rightarrow 1^-} \frac{\sqrt{u\sigma + 2C}}{B(C, u, \sigma, \rho)}$$

is bounded for any range of the parameters  $C, u$  and  $\sigma$  for which  $\sigma/C$  and  $1/u$  are bounded.

We continue the analysis begun in [1] of some examples linked with estimation of a signal in an additive white noise.

**EXAMPLE 1.** The set  $V$  consists of all functions on  $(0, 1)$  whose variation does not exceed  $M$  (see [1], example 2). We apply part (a) of Theorem 3 to the set  $V_1$  of functions in  $V$  orthogonal to the unit element. It is not hard to show that  $K(\theta)$  is finite only for step functions  $\theta \in V_1$  with variation exactly  $M$ ; for such functions  $K(\theta) = J(\theta) - 2$ , where  $J(\theta)$  is the number of steps. Part (a) of Theorem 3 yields the following proposition.

Let  $dX(t) = S(t) dt + \sigma dw(t)$ , where  $w$  is a Wiener process and  $S$  is a function of bounded variation;  $\text{Var}_{t \in (0, 1)} S(t) \leq M$ . Then the MLE  $\hat{S}$  for  $S$  in the indicated class of functions based on the observation  $X$  is a step function with probability, its variation is equal to  $M$ , and the random number  $J$  of steps of  $\hat{S}$  satisfies

$$E \exp(aJ) \leq \exp(\sqrt{\pi/8} e^a \sigma^{-1} M + 2a)$$

for any  $a \geq 0$ .

**EXAMPLE 2.** The set  $V$  consists of all increasing functions in  $L_2[0, 1]$ ; as in Example 3 in [1], we pass over to  $V_{a,b}$ . It is again clear that  $K(\theta)$  is finite only for step functions  $\theta$ ; from part (a) of Theorem 3 we conclude that  $\hat{\theta}$  is a step function with probability 1. However, after passing to the limit in  $a, b$  the number of steps becomes infinite (they cluster at the ends of the interval).

**REMARK 2.** Considering  $V$  to be the set of all  $L_2$  functions on the square  $[0, 1] \times [0, 1]$  that are increasing in both arguments, we conclude that the ML-estimation of such functions is impossible in additive two-dimensional white noise; indeed, even limiting ourselves to functions with two values 0 and 1, we obtain a compact set  $V$  not possessing the GB-property, as Dudley showed [5].

**EXAMPLE 3.** The set  $V$  consists of all functions  $\theta$  satisfying a Lipschitz condition  $|\theta(s) - \theta(t)| \leq M|s - t|$ ; see Example 1 for  $\alpha = 1$  in [1]. Part (a) of Theorem 3 applied to

the set  $V_1$  of functions in  $V$  orthogonal to the unit element makes it possible to draw only the following conclusion:  $d\hat{S}(t)/dt = \pm M$  for almost all  $t$  with probability 1. The fact is that any function in  $V_1$  (not only piecewise linear) with a derivative  $\pm M$  is an extreme point of  $V_1$ . In effect, with probability 1 the function  $\hat{S}$  is not piecewise linear; the intervals where it is linear make up the complement to a Cantor set.

REMARK 3. If so desired, Examples 1-2 can be generalized in the following direction. The condition  $\text{Var}_{t \in (0,1)} S(t) \leq M$  can be written in the form

$$\int_0^1 \left| \frac{d}{dt} S(t) \right| dt \leq M,$$

interpret as follows: the derivative of  $S$  is a finite mass in the sense of generalized function theory, and the norm of this mass does not exceed  $M$ . Similarly, we can examine the more general condition  $\int_0^1 |LS(t)| dt \leq M$ , where  $L$  is some linear differential operator. In that case the MLE is a piecewise smooth function, satisfying the equation  $LS(t) = 0$  on each piece. The condition that  $S$  is increasing can be generalized to the condition  $LS(t) \geq 0$ .

The proofs of Theorems 1-4 will be given in part 3 of this paper [2]. For convenience of orientation, we point out what topics comprise this paper (parts 2 and 3) and how these topics are interrelated.

Topics 1-3: the relationship between the probability and geometric considerations. Here there are three formally independent topics: (1) the thickness of an infinite-dimensional  $GB$ -compact set as the volume of the joint spectrum averaged with respect to Gaussian measure (Theorem 6, as well as Lemma 2); (2) geometric-probabilistic analyses for the case where  $V$  is a finite-dimensional polyhedron (Lemma 1; Theorem 5 is also relevant here); (3) the same where  $V$  is a finite-dimensional convex solid with smooth boundary ([2], Lemmas 1, 2). Topics 2 and 3 deal essentially with the same thing but in entirely different languages; it is easier to develop them independently rather than derive one from the other.

Topic 4: the distance of the MLE from the true value of the parameter ([2], Lemmas 3, 4) depends on topic 2.

Topics 5-7: probabilistic results on the behavior of the MLE. Here again there are three formally independent topics; all three rely on topic 4; moreover, topic 5 relies on topic 2, and topics 6 and 7 on topic 3. Topic 5: the probability that the MLE occurs on a face of given dimension if  $V$  is a finite-dimensional polyhedron (Theorem 1); topic 6: the same if  $V$  is infinite-dimensional (Theorem 3 and also Lemmas 5, 6 in [2]); topic 7: the distance between the MLE's for two strongly correlated sample points ([2], Theorems 2 and 4 and Lemmas 7, 8). It would appear that topic 6 ought to depend on topic 5 but it was found to be easier to develop it independently. In a certain sense topic 7 is also about the same thing, but in an entirely different language, in order to give a nontrivial result when  $V$  has no nontrivial faces.

Chevet [3] introduced a scale of geometric characteristics  $h_k$  for convex  $GB$ -compact sets in a Hilbert space;  $h_k(V)$  has come to be called the  $k$ -thickness of the set  $V$ . Starting out from the well-known integral cross-sectional measures  $W_k^*$ , she observed that under suitable numbering and norming they cease to depend on the dimension of the space containing the given finite-dimensional set; this allowed the  $h_k$  to be defined for finite-dimensional convex compact sets in Hilbert space; then an arbitrary convex compact set was approximated from within by finite-dimensional ones; here it turned out that  $h_k(V) < +\infty$  if and only if  $h_1(V) < +\infty$ , i.e., for  $V \in GB$ . A relationship was found between the  $k$ -thickness and the moments of the supremum of a Gaussian process. Finally, Chevet carried over to the infinite-dimensional case the classical inequalities of Fenchel-Alexandrov for mixed volumes. These inequalities are a basic tool for proving results of this paper. See a modern introduction to the theory of cross-sectional measures in [4], Chapter 4; Section 9.9 of the cited chapter considers the infinite-dimensional case following Chevet.

We give two new definitions for  $h_k$ ; in contrast to Chevet's definition, they are related directly to a Gaussian measure and do not use finite-dimensional approximation. The normalization of the thickness used by Chevet (and for  $k = 1$  by Sudakov) seems to us to be not the most convenient; it is more convenient to use the expression

$$(2) \quad \mathcal{M}_k(V) = (2\pi)^{-k/2} k! h_k(V), \quad k = 0, 1, \dots,$$

cf. [3], (3.6.2). We shall also call the quantity  $\mathcal{M}_k(V)$  the  $k$ -thickness of the set  $V$ . Inequalities (4.2.1) and (4.2.2) in [3] can now be written as

$$(3) \quad \mathcal{M}_{k-1}(V) \mathcal{M}_{k+1}(V) \leq \mathcal{M}_k^2(V),$$

$$(4) \quad \mathcal{M}_k(V) \leq \mathcal{M}_1^k(V), \quad k = 0, 1, \dots$$

For every  $GB$ -set  $V \subset E_0$  the collection of probability measures  $\{\gamma_\theta: \theta \in V\}$  has a least upper bound  $\gamma_V$  in the set of all (not only probability) finite measures on  $E$ ; it is clear that the measure  $\gamma_V$  has the following density with respect to the measure  $\gamma$ :

$$(5) \quad \frac{\gamma_V(dx)}{\gamma(dx)} = \sup_{\theta \in V} \mathcal{L}(\theta, x) = \exp \left( \sup_{\theta \in V} \left( \langle \theta, x \rangle - \frac{1}{2} \|\theta\|^2 \right) \right).$$

**Theorem 5.**  $\gamma_V(E) = \sum_{k=0}^{\infty} \mathcal{M}_k(V)/k!$

From (4) and this theorem we obtain

**Corollary 1.**  $\gamma_V(E) \leq \exp(\mathcal{M}_1(V))$ .

Noting that the functionals  $\mathcal{M}_k$  are homogeneous, we find that

$$\gamma_{aV}(E) = \sum_{k=0}^{\infty} \mathcal{M}_k(V) a^k / k!$$

for any  $a > 0$ , and hence we obtain an equivalent definition for  $k$ -thickness.

**Corollary 2.**

$$\begin{aligned} \mathcal{M}_k(V) &= \left. \frac{d^k}{da^k} \gamma_{aV}(E) \right|_{a=0+} \\ &= \left. \frac{d^k}{da^k} \left( \int \exp \left( \sup_{\theta \in V} \left( a \langle \theta, x \rangle - \frac{1}{2} a^2 \|\theta\|^2 \right) \right) \gamma(dx) \right) \right|_{a=0+}, \quad k = 0, 1, \dots \end{aligned}$$

One further equivalent definition of  $k$ -thickness will be given below in terms of the "joint spectrum" of several realizations of a random process. Here special care must be given to choosing a modification of the process. For  $GC$ -sets there is no problem since the realizations can be assumed to be continuous. However, in the general case the separable modification is found to be insufficient; the natural modification introduced in [6] is required. To every set  $V \subset E_0$ , a Gaussian random process  $(\theta, x)$  is defined where  $\theta$  ranges over  $V$  and  $x$  ranges over the space  $E$  equipped with the Gaussian measure  $\gamma$ . This process has a natural modification if and only if  $V \in GB\sigma$  (a countable union of  $GB$ -sets) [6]. It is understood below that  $(\theta, x)$  denotes precisely the natural modification of the above-mentioned process.

Let  $V \in GB$  (or  $GB\sigma$ ) and  $x_1, \dots, x_k \in E$ ; we call the set  $\text{spec}(x_1, \dots, x_k|V) = \{((\theta, x_1), \dots, (\theta, x_k)): \theta \in V\}$  lying in  $\mathbb{R}^k$  the joint spectrum for  $x_1, \dots, x_k$  on  $V$ . If  $V$  is a convex  $GB$ -compact set, then the joint spectrum is (a.s.) a convex bounded set; if  $V \in GC$ , it is closed and if  $V \notin GC$ , this is not necessarily so. We point out that a separable modification would not guarantee the convexity of the joint spectrum. The natural modification does guarantee it since the mapping  $\theta \rightarrow ((\theta, x_1), \dots, (\theta, x_k))$  is algebraically linear on the linear span of the set  $V$  even if it is not continuous on  $V$ .

**Theorem 6.** For any convex  $GB$ -compact set  $V \subset E_0$  and  $k = 0, 1, \dots$ ,

$$\mathcal{M}_k(V) = \frac{1}{\pi^k} \int \dots \int \text{mes}_k \text{spec}(x_1, \dots, x_k|V) \gamma(dx_1) \dots \gamma(dx_k);$$

here  $\text{mes}_k$  is Lebesgue measure in  $\mathbb{R}^k$ ;  $\pi_k = \pi^{k/2}/\Gamma(k/2 + 1)$  is the volume of the  $k$ -dimensional unit sphere.

**Lemma 1.** Let  $V$  be a finite-dimensional convex polyhedron in  $\mathbb{R}^n$ , and let  $F_k$  be defined as in Theorem 1. Then

$$\gamma_V\{x: \hat{\theta}(x) \in F_k\} = \mathcal{M}_k(V)/k!, \quad k = 0, 1, \dots, n.$$

**PROOF.** Consider a  $k$ -dimensional face  $F$  of the polyhedron  $V$  ( $F$  includes no points lying on faces of lower dimension), the  $k$ -dimensional subspace  $E_F \subset E_0$  parallel to it and its annihilator  $E_F^\perp = \{x \in E: \langle \eta, x \rangle = 0 \forall \eta \in E_F\}$ , as well as the affine subspace  $\tilde{F} = F + E_F$  containing  $F$ . In accordance with the decomposition  $E = \tilde{F} \oplus E_F^\perp$  each measure  $\gamma_\theta$  splits into the product of its projections  $\gamma_\theta = (\gamma_\theta|_{\tilde{F}}) \otimes (\gamma_\theta|_{E_F^\perp})$ . It is easy to see that for  $\theta \in F$  the set  $\{x - \hat{\theta}(x): x \in E, \hat{\theta}(x) = \theta\}$  is a convex cone  $K_F^0$  in  $E_F^\perp$  not depending on  $\theta$ . Therefore,  $\{x \in E: \hat{\theta}(x) \in F\} = F + K_F^0$ . The measure  $\gamma_V$  clearly coincides on this set with the measure

$$\gamma_F = \sup_{\theta \in F} \gamma_\theta = \sup_{\theta \in F} ((\gamma_\theta|_{\tilde{F}}) \otimes (\gamma_\theta|_{E_F^\perp})) = (\sup_{\theta \in F} (\gamma_\theta|_{\tilde{F}})) \otimes (\gamma_{\theta_0}|_{E_F^\perp}),$$

where  $\theta_0$  is the unique point in  $F \cap E_F^\perp$ ; however, the first factor coincides on the set  $F$  with the measure  $(2\pi)^{-k/2} \text{mes}_k$ . Thus

$$\gamma_V\{x \in E: \hat{\theta}(x) \in F\} = ((2\pi)^{-k/2} \text{mes}_k F)((\gamma|_{E_F^\perp})(K_F^0)) = (2\pi)^{-k/2} (\text{mes}_k F) \gamma(K_F),$$

where  $K_F = K_F^0 + E_F$ . Summing over all  $k$ -dimensional faces, we find that

$$\gamma_V\{x \in E: \hat{\theta}(x) \in F_k\} = (2\pi)^{-k/2} \sum_F (\text{mes}_k F) \gamma(K_F).$$

It remains to apply Lemma 3.5 in [3], according to which

$$h_k(V) = \sum_F (\text{mes}_k F) \gamma(K_F), \quad k = 0, 1, \dots, n.$$

**PROOF OF THEOREM 5.** If  $V$  is a finite-dimensional polyhedron  $\subset \mathbb{R}^n$ , applying Lemma 1, we find that

$$\gamma_V(E) = \sum_k \gamma_V\{x: \hat{\theta}(x) \in F_k\} = \sum_{k=0}^n \frac{1}{k!} \mathcal{M}_k(V).$$

In the general case we approximate  $V$  from within by finite-dimensional polyhedra  $V_n$ ,  $V_1 \subset V_2 \subset \dots \subset V$ , with  $\bigcup_{n=1}^\infty V_n$  dense in  $V$ ; here  $\mathcal{M}_k(V) = \lim_{n \rightarrow \infty} \mathcal{M}_k(V_n)$ ,  $k = 0, 1, \dots$ , according to Proposition 3.9 in [3], and

$$\gamma_V = \sup_{\theta \in V} \gamma_\theta = \sup_n \sup_{\theta \in V} \gamma_\theta = \sup_n \gamma_{V_n},$$

so

$$\gamma_V(E) = \lim_{n \rightarrow \infty} \gamma_{V_n}(E) = \lim_{n \rightarrow \infty} \sum_{k=0}^n \frac{1}{k!} \mathcal{M}_k(V_n) = \sum_{k=0}^\infty \frac{1}{k!} \mathcal{M}_k(V).$$

The following general property of a natural modification will be used in the proof of Theorem 6.

**Lemma 2.** Let  $\xi(\omega, t)$  be a natural modification of some random process,  $\omega \in \Omega$ ,  $t \in T$ , and let  $S$  be dense in  $T$  in the following sense: for any  $t \in T$  there are  $s_n \in S$ ,  $n = 1, 2, \dots$ , such that  $\xi(\omega, s_n) \rightarrow \xi(\omega, t)$ ,  $n \rightarrow \infty$ , for almost all  $\omega$  (the corresponding set of probability 1 may depend on  $t$ ). Then there is a set  $\Omega_1 \subset \Omega$  of probability 1 possessing the following property: for any  $t \in T$  there exist  $s'_n \in S$ ,  $n = 1, 2, \dots$ , such that  $\xi(\omega, s'_n) \rightarrow \xi(\omega, t)$ ,  $n \rightarrow \infty$ , for all  $\omega \in \Omega_1$ .

**PROOF.** By the definition of a natural modification, we can introduce a metric  $\rho$  on  $T$  such that  $(T, \rho)$  is a separable metric space and  $\xi(\omega, t)$  is continuous in  $t$  on  $(T, \rho)$  for all  $\omega$  in some  $\Omega_2 \subset \Omega$  of probability 1. Let  $\{t_m\}_{m=1}^\infty$  be a fixed countable set dense in  $(T, \rho)$ ;

for each  $m$ , fix  $s_{m,n} \in S$  such that

$$P\{|\xi(\omega, s_{m,n}) - \xi(\omega, t_m)| > 1/n\} \leq 2^{-m-n}.$$

The sum of these probabilities over all  $m$  and  $n$  is finite. Therefore there is a set  $\Omega_3 \subset \Omega$  of probability 1 possessing the following property:  $|\xi(\omega, s_{m,n}) - \xi(\omega, t_m)| \leq 1/n$  for all  $\omega \in \Omega_3$  and all pairs  $(m, n)$  except a finite (depending on  $\omega$ ) number of pairs. Put  $\Omega_1 = \Omega_2 \cap \Omega_3$ . Let  $t \in T$  be given. Choose  $m_1, m_2, \dots$ , such that  $t_{m_n} \rightarrow t$ ,  $n \rightarrow \infty$ , in  $(T, \rho)$ . Put  $s'_n = s_{m_n}$ . Then for large  $n$  we have

$$\begin{aligned} |\xi(\omega, s'_n) - \xi(\omega, t)| &\leq |\xi(\omega, s_{m_n}) - \xi(\omega, t_{m_n})| + |\xi(\omega, t_{m_n}) - \xi(\omega, t)| \\ &\leq 1/n + |\xi(\omega, t_{m_n}) - \xi(\omega, t)| \rightarrow 0, \quad n \rightarrow \infty, \end{aligned}$$

for all  $\omega \in \Omega_1$ , as required.

**Corollary 3.** Let  $V \subset E_0$  be a GB-set and let  $V_0 \subset V$ . If  $V_0$  is dense in  $V$ , then  $\text{spec}(x_1, \dots, x_k | V_0)$  is dense in  $\text{spec}(x_1, \dots, x_k | V)$  for almost all  $(x_1, \dots, x_k)$ .

Note that the assertion is trivial for GC-sets.

**PROOF OF THEOREM 6.** We first reduce the general case to the finite-dimensional one. To do this, we approximate  $V$  from within by finite-dimensional convex compact sets  $V_n$ ,  $n = 1, 2, \dots$ , where  $\mathcal{M}_k(V) = \lim_{n \rightarrow \infty} \mathcal{M}_k(V_n)$ , as already noted in the proof of Theorem 5. By Corollary 3 the union (over  $n$ ) of the convex sets  $\text{spec}(x_1, \dots, x_k | V_n)$  is dense in the convex set  $\text{spec}(x_1, \dots, x_k | V)$  a.s., and therefore

$$\begin{aligned} &\int \cdots \int \text{mes}_k \text{spec}(x_1, \dots, x_k | V) \gamma(dx_1) \cdots \gamma(dx_k) \\ &= \int \cdots \int \lim_{n \rightarrow \infty} \text{mes}_k \text{spec}(x_1, \dots, x_k | V_n) \gamma(dx_1) \cdots \gamma(dx_k) \\ &= \lim_{n \rightarrow \infty} \int \cdots \int \text{mes}_k \text{spec}(x_1, \dots, x_k | V_n) \gamma(dx_1) \cdots \gamma(dx_k). \end{aligned}$$

Hence, it suffices to prove Theorem 6 for the case where  $V$  is finite-dimensional. The space  $E$  can then also be assumed to be finite-dimensional of dimension  $n$ . The set  $\text{spec}(x_1, \dots, x_k | V)$  is the image of  $V$  under the linear mapping  $A_x: E \rightarrow \mathbb{R}^k$  which is defined by  $A_x(\theta) = (\langle \theta, x_1 \rangle, \dots, \langle \theta, x_k \rangle)$ ; here and elsewhere  $x = (x_1, \dots, x_k)$ . Choose a fixed orthonormal basis in  $E$ ; the matrix of the mapping  $A_x$  in this basis (denote it also by  $A_x$ ) has  $k$  rows and  $n$  columns. Each element of this matrix depends on the point  $x$  of the probability space  $(E \oplus \cdots \oplus E, \gamma \otimes \cdots \otimes \gamma)$  and in this sense can be viewed as a random variable. It is easy to see that all the elements of  $A_x$  are independent standard Gaussian random variables. Orthogonalizing the rows of  $A_x$ , we obtain a factorization  $A_x = B_x C_x$ , where  $B_x$  is a lower triangular square matrix of order  $k$  and  $C_x$  is a matrix of  $k$  orthonormal rows of  $n$  elements. Examining the orthogonalization process, we can easily show that all the elements of  $B_x$  are independent, all the off-diagonal elements have a standard normal distribution and all diagonal elements have the following distributions:  $\chi_n, \chi_{n-1}, \dots, \chi_{n-k+1}$  ( $\chi_l$  is the distribution of the square root of the sum of the squares of  $l$  independent standard normal variables). Therefore,

$$\begin{aligned} \mathbf{E} \det B_x &= (\mathbf{E} \chi_n) (\mathbf{E} \chi_{n-1}) \cdots (\mathbf{E} \chi_{n-k+1}) \\ &= \frac{\sqrt{2} \Gamma\left(\frac{n+1}{2}\right)}{\Gamma\left(\frac{n}{2}\right)} \frac{\sqrt{2} \Gamma\left(\frac{n}{2}\right)}{\Gamma\left(\frac{n-1}{2}\right)} \cdots \frac{\sqrt{2} \Gamma\left(\frac{n-k+2}{2}\right)}{\Gamma\left(\frac{n-k+1}{2}\right)} = 2^{k/2} \frac{\Gamma\left(\frac{n+1}{2}\right)}{\Gamma\left(\frac{n-k+1}{2}\right)}. \end{aligned}$$

Further, it is not hard to show that  $B_x$  and  $C_x$  are independent and that the matrix  $C_x$  has a unique probability distribution invariant under rotation or more precisely under right-multiplication by any orthogonal matrix of order  $n$ ; it is essentially a question of a

uniform distribution on the Stiefel manifold  $V_k(\mathbb{R}^n)$ , and just below of a uniform distribution on the Grassman manifold  $G_k(\mathbb{R}^n)$ . We have

$$\begin{aligned} \mathbf{E} \operatorname{mes}_k \operatorname{spec}(x|V) &= \mathbf{E} \operatorname{mes}_k B_x C_x V \\ &= (\mathbf{E} \det B_x)(\mathbf{E} \operatorname{mes}_k C_x V) \\ &= 2^{k/2} \frac{\Gamma((n+1)/2)}{\Gamma((n-k+1)/2)} \mathbf{E} \operatorname{mes}_k C_x V. \end{aligned}$$

The set  $C_x V$  is isometric to the orthogonal projection of  $V$  on the  $k$ -dimensional subspace spanned by the rows of the matrix  $C_x$ . Therefore,  $\mathbf{E} \operatorname{mes}_k C_x V$  is the mean volume of the  $k$ -dimensional projection of  $V$ . We use a consequence of Kubota's formula [3], formula 4.4.4"; it says that the mean volume of the  $k$ -dimensional projection of  $V$  is equal to

$$\Gamma\left(\frac{n}{2}+1\right) \left[ \binom{n}{k} \Gamma\left(\frac{k}{2}+1\right) \Gamma\left(\frac{n-k}{2}+1\right) \right]^{-1} h_k(V);$$

by means of the duplication formula for the  $\Gamma$ -function

$$\Gamma(2z) = \pi^{-1/2} 2^{2z-1} \Gamma(z) \Gamma(z + \frac{1}{2})$$

we now find that

$$\begin{aligned} \frac{1}{\pi_k} \mathbf{E} \operatorname{mes}_k \operatorname{spec}(x|V) &= \frac{\Gamma(k/2+1)}{\pi^{k/2}} 2^{k/2} \frac{\Gamma((n+1)/2)}{\Gamma((n-k+1)/2)} \frac{k!(n-k)!}{n!} \\ &\cdot \frac{\Gamma(n/2+1)}{\Gamma(k/2+1)\Gamma((n-k)/2+1)} \frac{(2\pi)^{k/2}}{k!} \mathcal{M}_k(V) \\ &= 2^k \frac{\Gamma((n+1)/2)\Gamma(n/2+1)}{\Gamma(n+1)} \\ &\cdot \frac{\Gamma(n-k+1)}{\Gamma((n-k+1)/2)\Gamma((n-k)/2+1)} \mathcal{M}_k(V) \\ &= 2^k \frac{\sqrt{\pi}}{2^n} \frac{2^{n-k}}{\sqrt{\pi}} \mathcal{M}_k(V) = \mathcal{M}_k(V). \end{aligned}$$

REMARK 4. Theorem 6 can be proved differently for the finite-dimensional case by introducing a large number of "extraneous" measurements, as was done in the proof of Theorem 3 in [7].

Received by the editors  
December 9, 1983

REFERENCES

- [1] B. S. TSIREL'SON, *A geometric approach to maximum likelihood estimation for infinite-dimensional Gaussian location. I*, Theory Prob. Appl., 27 (1982), pp. 411-418.
- [2] B. S. TSIREL'SON, *A geometric approach to maximum likelihood estimation for infinite-dimensional Gaussian location. III*, Theory Prob. Appl., (in press).
- [3] S. CHEVET, *Processus gaussiens et volumes mixtes*, Z. Wahrsch Verw. Gebiete, 36 (1976), pp. 47-65.
- [4] YU. D. BURAGO AND V. A. ZALGALLER, *Geometric Inequalities*, Nauka, Leningrad, 1980. (In Russian.)
- [5] R. M. DUDLEY, *Lower layers in  $\mathbb{R}^2$  and convex sets in  $\mathbb{R}^3$  are not GB classes*, Lect. Notes Math., 709 (1979), pp. 97-102.
- [6] B. S. TSIREL'SON, *A natural modification of a random process and its applications to random series of functions and Gaussian measures*, Zap. Nauchr. Sem. Leningrad Otdel. Mat. Inst. Steklon, 55 (1976), pp. 35-63. (In Russian.)



- [7] V. N. SUDAKOV AND B. S. TSIREL'SON, *Extremal properties of halfspaces for spherically invariant measures*, Zap. Nauchr. Sem. Leningrad Otdel. Mat. Inst. Steklon, 41 (1974), pp. 14-24. (In Russian.)

## THE UNIFORM DISTRIBUTION ON COMPACT HOMOGENEOUS SPACES AND THE KANTOROVICH-RUBINSHTEIN METRIC

V. A. KAIMANOVICH

(Translated by Yona Ellis)

Let  $X$  be a compact  $G$ -space with a Haar measure  $m$  (e.g.,  $G$  is a compact group,  $X = G$ ). A number of papers (see [1]-[4]) have studied the problem of the *uniform distribution* on  $X$ , i.e., the weak convergence of the measures  $\mu_1 \cdots \mu_n \nu$  or  $\mu^n \nu$  to  $m$  ( $\mu_i$  and  $\nu$  are probability measures on  $G$  and  $X$ , respectively and  $\mu^n$  is the  $n$ -fold convolution of  $\mu$ ). Below we apply a new technique to the investigation of this problem based on the metrization of the weak topology by the Kantorovich-Rubinshtein (KR) metric and we generalize old results of Arnol'd-Krylov and Ullrich-Urbanik.

### 1. The KR Metric

Let  $(X, \rho)$  be a compact metric space,  $M(X)$  the space of Borel probability measures on  $X$  and  $\text{Lip}(X, \rho)$  the space of functions  $u: X \rightarrow \mathbb{R}$  such that  $u(s) - u(t) \leq \rho(s, t) (\forall s, t \in X)$ . The KR-metric

$$d(\nu_1, \nu_2) = \inf \int \rho(s, t) d\Psi(s, t),$$

where the inf is taken over all measures  $\Psi \in M(X \times X)$  whose projections onto the first and second factors are  $\nu_1$  and  $\nu_2$ , metrizes the weak topology in  $M(X)$ , and the inf is always attained for a measure  $\Psi$  if and only if  $u(s) - u(t) = \rho(s, t) (\forall (s, t) \in \text{supp } \Psi)$  for some  $u \in \text{Lip}(X, \rho)$ . The dual definition is  $d(\nu_1, \nu_2) = \sup \{ \langle f, \nu_1 - \nu_2 \rangle : f \in \text{Lip}(X, \rho) \}$  [5].

Everywhere below (except in Section 6) we assume that the *topological group  $G$  acts continuously and minimally on  $X$  by isometries*, i.e.,  $G$  is embedded in the compact group  $\text{Iso}(X)$  of all isometries of  $X$  equipped with the invariant metric  $\rho^*(g_1, g_2) = \sup_x \rho(g_1 x, g_2 x)$ . Obviously,  $d(g\nu_1, g\nu_2) = d(\nu_1, \nu_2)$ ,  $d(\mu\nu_1, \mu\nu_2) \leq d(\nu_1, \nu_2)$ . We call a measure  $\mu$  on  $G$  *transitive on  $X$*  if  $\rho(s, \text{supp } \mu^n \delta_t) \rightarrow 0$ ,  $n \rightarrow \infty (\forall s, t \in X)$ . For  $X = G$  this is equivalent to  $\mu$  being *strictly aperiodic* [1].

**Lemma 1.** *If  $d(\mu\nu_1, \mu\nu_2) = d(\nu_1, \nu_2)$  and  $\text{supp } \mu' \subset \text{supp } \mu$ , then  $d(\mu'\nu_1, \mu'\nu_2) = d(\nu_1, \nu_2)$ .*

**Lemma 2.** *If the measure  $\mu$  is transitive on  $X$  and  $\nu_1 \neq \nu_2 \in M(X)$ , then  $d(\mu^n \nu_1, \mu^n \nu_2) < d(\nu_1, \nu_2)$  for some  $n$  (depending on  $\nu_{1,2}$ ).*

**PROOF.** Let  $\Psi$  be a fixed measure realizing  $d(\nu_1, \nu_2)$ , and  $(s, t) \in \text{supp } \Psi$ ,  $\rho(s, t) = \varepsilon > 0$ . If  $d(\mu^n \nu_1, \mu^n \nu_2) = d(\nu_1, \nu_2)$ , then the measure  $\mu^n \Psi$  realizes  $d(\mu^n \nu_1, \mu^n \nu_2)$ , i.e., there is a  $u_n \in \text{Lip}(X, \rho)$  such that

$$u_n(gs) - u_n(gt) = \rho(gs, gt) = \varepsilon (\forall g \in \text{supp } \mu^n).$$

The values of  $u_n(gs)$  for  $g \in \text{supp } \mu^n$  for large  $n$  can be arbitrarily close to the smallest value of  $u_n$  on  $X$ , whence  $\varepsilon = 0$  and  $\nu_1 = \nu_2$ .