

## A MULTI-PARAMETER GAUSSIAN PROCESS<sup>1</sup>

BY WON JOON PARK

Wright State University

**1. Introduction.** Let  $(\Omega, \mathcal{F}, P)$  be a probability space and let  $A$  be the  $p$ -dimensional unit rectangle ( $p \geq 2$ ). We denote by  $(A, \mathcal{A}, \mu_p)$  the ordinary Lebesgue measure space. Let  $\{X(u, \omega): u \in A\}$  be a Gaussian process defined on  $(\Omega, \mathcal{F}, P)$  with the properties:

$$(1.1) \quad X(u, \omega) = 0 \quad \text{a.s. for every } u \text{ in } A_0 \text{ where}$$

$$A_0 = \{(u_1, \dots, u_p) \in A: u_j = 0 \text{ for some } j \text{ with } 1 \leq j \leq p\}.$$

$$(1.2) \quad E[X(u, \omega)] = \int_{\Omega} X(u, \omega) dP(\omega) = 0 \quad \text{for every } u \text{ in } A.$$

$$(1.3) \quad E[X(u, \omega)X(v, \omega)] = \min(u_1, v_1) \cdots \min(u_p, v_p) = R(u, v) \\ \text{for every } u = (u_1, \dots, u_p) \text{ and } v = (v_1, \dots, v_p) \text{ in } A.$$

By considering an expansion in terms of Haar functions on  $A$ , it is shown that  $X(u, \omega)$  can be realized in the space  $C(A)$  of real continuous functions on  $A$  which vanish at  $A_0$ , i.e.

$$(1.4) \quad \text{Almost all sample functions of } X(u, \omega) \text{ are continuous.}$$

For  $p = 2$ , the existence of the above Gaussian process  $X(u, \omega)$  is shown by Yeh [15] and Kuelbs [10]. We will call a Gaussian process  $X(u, \omega)$  with the properties (1.1)–(1.4) the  $p$ -parameter Gaussian process. We then examine the interrelationship between the  $p$ -parameter Gaussian process and its reproducing kernel Hilbert space  $H(R)$ . Let  $L^2(A)$  denote the space of Lebesgue square-integrable functions on  $A$  with an inner product  $(f, g) = \int_A f(u)g(u) d\mu_p(u)$  and norm  $\| \cdot \|$ . We also define a stochastic integral  $I(f) = \int_A f(u) dX(u, \omega)$  for  $f \in L^2(A)$  with respect to the  $p$ -parameter Gaussian process in two different ways and show that they are identical. From this we show that the  $p$ -parameter Gaussian process has an a.s. uniformly convergent orthonormal expansion.

Defining a Gaussian random set function by

$$(1.5) \quad X(F, \omega) = \int_A 1_F(t) dX(t, \omega)$$

where  $F \in \mathcal{A}$  and  $1_F$  is the indicator function of  $F$ , we define the multiple Wiener integral (see Itô [6]) and show that any  $L^2$ -functional of the process has an orthogonal representation.

By appealing to the results obtained by Parzen [13], Kallianpur [9] and Oodaira [11], we can simply deduce the results: a translation theorem, equivalence of

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Gaussian measures and a zero-one law for the  $r$ -module of the  $p$ -parameter Gaussian process, which are generalizations of some known results due to Park [12], Shepp [14] and Yeh [16].

**2. Haar functions on  $L^2(A)$  and construction of the process.** Let  $\{\bar{g}_{n,j}\}$  denote the Haar functions on  $L^2(I)$ , where  $I = [0, 1]$ , i.e.

$$\begin{aligned}
 \bar{g}_{0,0} &\equiv 1 \\
 (2.1) \quad \bar{g}_{n,j}(s) &= 2^{(n-1)/2} && \text{if } s \in [j/2^{n-1}, j + \frac{1}{2}/2^{n-1}), \\
 &= -2^{(n-1)/2} && \text{if } s \in [j + \frac{1}{2}/2^{n-1}, j + 1/2^{n-1}), \\
 &= 0 && \text{otherwise;}
 \end{aligned}$$

for  $n = 1, 2, \dots$  and  $j = 0, 1, \dots, 2^{n-1} - 1$ .

It is shown by Haar [5] that  $\{\bar{g}_{n,j}\}$  is a complete orthonormal system (C.O.N.S.) in  $L^2(I)$ . We shall use the following notations:

$$u = (u_1, \dots, u_p) \text{ for } u \in A.$$

$$1_u = \text{the indicator function of } [0, u_1] \times \dots \times [0, u_p].$$

$D =$  set of all  $p$ -tuples  $n = (n_1, \dots, n_p)$  with nonnegative integers  $n_i (i = 1, 2, \dots, p)$ .

$$|n| = n_1 + \dots + n_p \text{ for } n = (n_1, \dots, n_p) \text{ in } D.$$

$$S_n = \{j = (j_1, \dots, j_p) : 0 \leq j_i \leq 2^{n_i-1} - 1, i = 1, 2, \dots, p\} \text{ for a fixed } n = (n_1, \dots, n_p) \text{ in } D.$$

For  $u \in A, n \in D$  and  $j \in S_n$ , define

$$(2.2) \quad g_{n,j}(u) = \bar{g}_{n_1,j_1}(u_1) \cdots \bar{g}_{n_p,j_p}(u_p)$$

where  $\{\bar{g}_{n,j}\}$  are Haar functions on  $L^2(I)$ . Then it is easy to see that  $\{g_{n,j}\} (n \in D, j \in S_n)$  is a C.O.N.S. in  $L^2(A)$ .

For each  $n \in D$  and  $j \in S_n$ , let us define

$$(2.3) \quad G_{n,j}(u) = (1_u, g_{n,j}) \quad (u \in A)$$

where  $g_{n,j}$  is given by (2.2). It follows clearly from the definition that  $G_{n,j}$  is a continuous function on  $A$ .

Now we shall define a multi-parameter Gaussian process. Let  $\{y_{n,j}\} (n \in D, j \in S_n)$  be mutually independent  $N(0, 1)$  random variables defined on  $(\Omega, \mathcal{F}, P)$ , where  $N(m, \sigma^2)$  denotes the normal distribution with mean  $m$  and variance  $\sigma^2$ .

Consider the series

$$(2.4) \quad \sum_{N=0}^{\infty} \sum_{|n|=N} \sum_{j \in S_n} y_{n,j}(\omega) G_{n,j}(u) \quad (\omega \in \Omega, u \in A)$$

where  $\sum_{|n|=N}$  means summing over all possible  $n \in D$  with  $|n| = N$ .

**THEOREM 1.** *The series (2.4) converges uniformly in  $u \in A$  with probability one.*

**PROOF.** Let

$$(2.5) \quad f_N(u, \omega) = \sum_{|n|=N} \sum_{j \in S_n} y_{n,j}(\omega) G_{n,j}(u) \quad \text{and}$$

$$(2.6) \quad Y_N(\omega) = \max_{|n|=N, j \in S_n} |y_{n,j}(\omega)|$$

for  $N = 0, 1, 2, \dots$ ; then

$$(2.7) \quad \max_{u \in A} |f_N(u, \omega)| \leq Y_N(\omega) \binom{N+p-1}{p-1} 2^{-(N+p)/2},$$

since for a fixed  $n \in D$ ,  $G_{n,j}$  are non-overlapping for different  $j \in S_n$ ,  $\max_{u \in A} |G_{n,j}(u)| \leq 2^{-(|n|+p)/2}$  and the number of terms in the summation  $\sum_{|n|=N}$  is  $\binom{N+p-1}{p-1}$ .

Let  $a_N > 0$ . Then since  $y_{n,j}$  are mutually independent  $N(0, 1)$  random variables,

$$(2.8) \quad \begin{aligned} P\{\omega : Y_N(\omega) > a_N\} &\leq \binom{N+p-1}{p-1} 2^{N-p} (2\pi)^{-\frac{1}{2}} 2 \int_{a_N}^{\infty} \exp(-s^2/2) ds \\ &\leq \binom{N+p-1}{p-1} 2^{N-p} (2\pi)^{-\frac{1}{2}} 2 a_N^{-(2p+2)} \\ &\quad \cdot \exp(-a_N^2/4) \int_{a_N}^{\infty} s^{2p+2} \exp(-s^2/4) ds. \end{aligned}$$

Choosing  $a_N = 2[(N+p-1) \ln 2]^{\frac{1}{2}}$ , it is easy to check that

$$(2.9) \quad P\{\omega : Y_N(\omega) > 2[(N+p-1) \ln 2]^{\frac{1}{2}}\} \geq C_p (N+p-1)^{-2},$$

where  $C_p$  is a finite constant which does not depend on  $N$ . Therefore

$$\sum_{N=0}^{\infty} P\{\omega : Y_N(\omega) > 2[(N+p-1) \ln 2]^{\frac{1}{2}}\} < \infty$$

hence by the Borel–Cantelli lemma,

$$(2.10) \quad P\{\omega : Y_N(\omega) > 2[(N+p-1) \ln 2]^{\frac{1}{2}}, \text{ infinitely often}\} = 0$$

i.e.,

$$(2.11) \quad P\{\omega : Y_N(\omega) \leq 2[(N+p-1) \ln 2]^{\frac{1}{2}}, \text{ for all } N \geq N_0(\omega)\} = 1.$$

Now  $\sum_{N=N_0(\omega)} \max_{u \in A} |f_N(u, \omega)| < \infty$  follows from (2.7) and (2.11). Consequently  $\sum_{N=0}^{\infty} \max_{u \in A} |f_N(u, \omega)|$  converges a.s., i.e. the series (2.4) converges uniformly in  $u \in A$  with probability one. This completes the proof of the theorem.

The above result is a generalization of a similar theorem proved by Ciesielski [3] for the standard Wiener process.

Let  $\Omega^*$  denote the set of  $\omega$  in  $\Omega$  such that the series (2.4) converges uniformly in  $u \in A$ . We note that  $P(\Omega^*) = 1$ . We now define a stochastic process

$$(2.12) \quad \begin{aligned} X(u, \omega) &= \sum_{N=0}^{\infty} \sum_{|n|=N} \sum_{j \in S_n} y_{n,j}(\omega) G_{n,j}(u) && (u \in A) \\ & && \text{if } \omega \in \Omega^* \end{aligned}$$

$$X(u, \omega) = 0 \quad \text{for every } u \in A \quad \text{if } \omega \notin \Omega^*.$$

The following corollary is obvious.

**COROLLARY 1.** *All sample functions of the stochastic process  $\{X(u, \omega): u \in A\}$  defined by (2.12) are continuous.*

**THEOREM 2.** *The stochastic process  $\{X(u, \omega): u \in A\}$  defined by (2.12) is Gaussian with mean function zero and covariance function*

$$(2.13) \quad R(u, v) = \min(u_1, v_1) \cdots \min(u_p, v_p) \quad (u, v \in A)$$

and satisfies

$$(2.14) \quad X(u, \omega) = 0 \quad \text{a.s. for every } u \in A_0.$$

**PROOF.** It follows immediately from the definition of the process that  $\{X(u, \omega): u \in A\}$  is a Gaussian process with mean function zero, and (2.14) holds. Denote the process  $X(u, \omega)$  simply as

$$(2.15) \quad X(u, \omega) = \sum_{j=0}^{\infty} y_j(\omega) G_j(u) \quad \text{a.s.}$$

Since  $y_j(\omega) G_j(u)$  are mutually independent  $N(0, G_j^2(u))$  random variables and  $\sum_{j=0}^{\infty} G_j^2(u) < \infty$ , we have

$$E[X(u, \omega) - \sum_{j=0}^n y_j(\omega) G_j(u)]^2 \rightarrow 0 \quad \text{as } n \rightarrow \infty.$$

Now it is easy to see that

$$(2.16) \quad E[X(u, \omega)X(v, \omega)] = \lim_{n \rightarrow \infty} E[\{\sum_{j=0}^n y_j(\omega) G_j(u)\} \{\sum_{j=0}^n y_j(\omega) G_j(v)\}].$$

By identifying the right-hand side (r.h.s) of (2.16) to

$$\begin{aligned} \text{iim}_{n \rightarrow \infty} \sum_{j=0}^n G_j(u) G_j(v) &= \sum_{j=0}^{\infty} (1_u, g_j)(1_v, g_j) \\ &= (1_u, 1_v) = \min(u_1, v_1) \cdots \min(u_p, v_p), \end{aligned}$$

we obtain (2.13) and thus the proof is complete.

It should be pointed out that Theorem 2 and Corollary 1 together imply the existence of the  $p$ -parameter Gaussian process defined in the introduction.

**3. The realization of the process on  $C(A)$ .** Here we shall relate the results obtained in the previous section to those obtained by Yeh [15].

Let  $(C(I^2), \underline{B}(C), m_W)$  be the Wiener space given in [15], where  $C(I^2)$  is the space of real-valued continuous functions on  $I^2$  which vanish on  $\{(s, t): s = 0 \text{ or } t = 0\}$ ,  $\underline{B}(C)$  is a  $\sigma$ -field generated by the cylinder sets in  $C(I^2)$  and  $m_W$  is the Wiener measure on  $C(I^2)$ . This measure space is also constructed by Kuelbs [10] by using different techniques—mainly Prokhorov's weak convergence of measures. The existence of the probability space  $(C(I^2), \underline{B}(C), m_W)$  was also shown by N. N. Cencov (*Akademica Nauk SSSR Doklady* **106**, 1956).

The triplet  $(C(I^2), \underline{B}(C), m_W)$  can be considered as a stochastic process where the random variables of the process are given by the coordinate variables  $x(s, t)$ ,  $x \in C(I^2)$ .

**THEOREM 3.** *The Gaussian process  $(C(I^2), \underline{B}(C), m_w)$  has mean function zero and covariance function*

$$(3.1) \quad \begin{aligned} R((s_1, t_1), (s_2, t_2)) &= E_{m_w}[x(s_1, t_1)x(s_2, t_2)] \\ &= \min(s_1, t_1) \min(s_2, t_2) \end{aligned}$$

for  $(s_1, t_1)$  and  $(s_2, t_2)$  in  $I^2$ .

**PROOF.** It is easy to check that the mean function is zero. (3.1) follows from the evaluation of the following integral: For  $0 < s_1 < s_2$  and  $0 < t_1 < t_2$ ,

$$\begin{aligned} &\int_{C(I^2)} x(s_1, t_1)x(s_2, t_2) dm_w(x) \\ &= \{(2\pi)^4 [s_1(s_2 - s_1)]^2 [t_1(t_2 - t_1)]^2\}^{-\frac{1}{2}} \int_{-\infty}^{\infty} (4) \int_{-\infty}^{\infty} u_{11} u_{22} \\ &\quad \cdot \exp\left\{-\frac{1}{2} [u_{11}^2/s_1 t_1 + (u_{12} - u_{11})^2/s_1(t_2 - t_1) \right. \\ &\quad \left. + (u_{21} - u_{11})^2/(s_2 - s_1)t_1 + (u_{22} - u_{21} - u_{12} + u_{11})^2/(s_2 - s_1)(t_2 - t_1)]\right\} \\ &\quad \cdot du_{11} du_{12} du_{21} du_{22} = s_1 t_1. \end{aligned}$$

Thus the theorem is proved.

Let  $\{X(u, \omega) : u \in A\}$  be the  $p$ -parameter Gaussian process and let  $B(X(u) : u \in A)$  be the  $\sigma$ -field generated by the process  $X(u, \omega)$  and  $\underline{B}(X(u) : u \in A)$  or simply  $\underline{B}(X)$  be the completion of  $B(X(u) : u \in A)$  under  $P$ . We usually replace the  $\sigma$ -field  $\mathcal{F}$  by  $\underline{B}(X)$  and consider  $(\Omega, \underline{B}(X), P, \{X(u) : u \in A\})$  to be the  $p$ -parameter Gaussian process. We shall write  $L^2(\Omega)$  for  $L^2(\Omega, \underline{B}(X), P)$ , the Hilbert space of  $\underline{B}(X)$ -measurable, real-valued functions square integrable with respect to  $P$  with an inner product

$$(x, y)_{L^2(\Omega)} = \int_{\Omega} x(\omega)y(\omega) dP(\omega)$$

and norm  $\| \cdot \|_{L^2(\Omega)}$ . From now if it is obvious that  $X(u, \omega)$  is a random variable then we may write  $X(u)$  instead of  $X(u, \omega)$ .

Let  $B(C(A))$  or simply  $B(C)$  denote the  $\sigma$ -field generated by all cylinder sets in  $C(A)$ . Let us define a map  $S : \Omega \rightarrow C(A)$  by

$$(3.2) \quad S(\omega) = X(\cdot, \omega) \quad \text{for } \omega \in \Omega$$

and also define a probability measure  $m_w$  on  $(C(A), B(C))$  by

$$(3.3) \quad m_w(E) = P\{S^{-1}(E)\} \quad \text{for } E \in B(C).$$

Then  $m_w$  has the following property for a cylinder set in  $C(A)$ :

$$(3.4) \quad m_w\{x \in C(A) : [x(u_1), \dots, x(u_n)] \in F\} = P\{\omega \in \Omega : [X(u_1, \omega), \dots, X(u_n, \omega)] \in F\}$$

where  $F$  is a Borel set in  $n$ -dimensional Euclidean space  $R^n$  and  $u_1, \dots, u_n$  are in  $A$ . The probability space  $(C(A), \underline{B}(C), m_w)$ , where  $\underline{B}(C)$  is the completion of  $B(C)$  under  $m_w$ , is a generalization of  $(C(I^2), \underline{B}(C), m_w)$  given by Yeh, since the mean function being zero and the covariance function determines a Gaussian measure

uniquely and when  $p = 2$  ( $C(A)$ ,  $B(C)$ ,  $m_w$ ) has the same covariance function as Yeh's.

**4. A stochastic integral with respect to  $X(u, \omega)$ .** Let  $X(u, \omega)$  be the  $p$ -parameter Gaussian process. We shall use the following notations:

(4.1) For  $u = (u_1, \dots, u_p)$  and  $v = (v_1, \dots, v_p)$  in  $A$ ,  
 $u < v$  mean  $u_i < v_i$  for  $i = 1, 2, \dots, p$ .

(4.2)  $\Delta_{u,v} = \prod_{i=1}^p [u_i, v_i]$  for  $u, v \in A$  with  $u < v$ .

(4.3)  $V(u, v, k)$  denotes the set of  $p$ -tuples  $s = (s_1, \dots, s_p)$  such that each  $s_i$  is either  $u_i$  or  $v_i$  and exactly  $k$  of  $s_i$  are  $u_i$  for  $u = (u_1, \dots, u_p)$  and  $v = (v_1, \dots, v_p)$ .

(4.4)  $\Delta_{u,v} X(\omega) = \sum_{k=0}^p (-1)^k \sum_{s \in V(u,v,k)} X(s, \omega)$  ( $u < v$ ).

(4.5)  $\bar{V}(n, j, k) = V(j + \frac{1}{2}/2^{n-1}, j + 1/2^{n-1}, k)$  for  $n \in D$  and  $j \in S_n$ ,

where  $V$  in the r.h.s. is given by (4.3) and,

for  $n = (n_1, \dots, n_p)$  and  $j = (j_1, \dots, j_p)$ ,

$j + \frac{1}{2}/2^{n-1} = (j_1 + \frac{1}{2}/2^{n_1-1}, \dots, j_p + \frac{1}{2}/2^{n_p-1})$  and

$j + 1/2^{n-1} = (j_1 + 1/2^{n_1-1}, \dots, j_p + 1/2^{n_p-1})$ .

(4.6)  $\bar{V}(n, j) = \bigcup_{k=0}^p \bar{V}(n, j, k)$  for  $n \in D$  and  $j \in S_n$ .

(4.7) For  $s \in \bar{V}(n, j)$  with  $s = (r_1/2^{n_1-1}, \dots, r_p/2^{n_p-1})$   
 $s - \frac{1}{2}$  means  $(r_1 - \frac{1}{2}/2^{n_1-1}, \dots, r_p - \frac{1}{2}/2^{n_p-1})$ .

For any  $n \in D$  let  $0 = u_i^0 < u_i^1 < \dots < u_i^{n_i} \leq 1$  ( $i = 1, \dots, p$ ) be a partition of  $A$ .

(4.8)  $S_n^* = \{j = (j_1, \dots, j_p) : 1 \leq j_i \leq n_i \text{ for } i = 1, 2, \dots, p\}$ .

(4.9)  $A_n = \{u^j = (u_1^{j_1}, \dots, u_p^{j_p}) \in A : j = (j_1, \dots, j_p) \in S_n^*\}$ .

(4.10) If  $j = (j_1, \dots, j_p)$  in  $S_n^*$ , then  $j-1 = (j_1-1, \dots, j_p-1)$ .

(4.11)  $\bar{y}_{n,j}(\omega) = 2^{(|n|-p)/2} \sum_{k=0}^p (-1)^k \sum_{s \in \bar{V}(n,j,k)} \Delta_{s-\frac{1}{2},s} X(\omega)$

for  $n \in D$  and  $j \in S_n$ .

LEMMA 1. For  $u, v, s$  and  $t$  in  $A$  with  $u < v$  and  $s < t$

$$E[\Delta_{u,v} X(\omega) \cdot \Delta_{s,t} X(\omega)] = \mu_p(\Delta_{u,v} \cap \Delta_{s,t}).$$

LEMMA 2.  $\{\bar{y}_{n,j}(\omega)\}$  ( $n \in D, j \in S_n$ ) are  $B(X)$ -measurable and mutually independent  $N(0, 1)$  random variables.

The proofs of Lemma 1 and Lemma 2 are purely computational and they are omitted here. Now we shall define a stochastic integral with respect to the  $p$ -parameter Gaussian process  $X(u, \omega)$ , denoted by  $I(f) = \int_A f(u) dX(u)$  for  $f \in L^2(A)$ .

We first define  $I(f)$  when  $f$  is a step function. If

$$(4.12) \quad f(u) = c_j \quad \text{for } u \in \Delta_{u^{j-1}, u^j} \quad (u^j \in A_n)$$

where  $\Delta_{u^{j-1}, u^j}$  and a partition  $A_n$  are given by (4.2) and (4.9), then we define

$$(4.13) \quad I(f) = \int_A f(u) dX(u) = \sum_{j \in S_n^*} c_j \Delta_{u^{j-1}, u^j} X(\omega)$$

where  $S_n^*$  is given in (4.8).

In fact, we shall accept as  $I(f)$  any random variable equal almost surely to the sum on the right. As defined in (4.13) the integral is determined uniquely by  $f$ , neglecting  $I(f)$  values on a set of zero probability. For each step function  $f$ ,  $I(f)$  is clearly  $B(X)$ -measurable. Let  $g$  be a step function of the same type as (4.12).

LEMMA 3. *The stochastic integral satisfies:*

$$(4.14) \quad I(af + bg) = aI(f) + bI(g)$$

for real numbers  $a$  and  $b$ ;

$$(4.15) \quad E[I(f) \cdot I(g)] = (f, g),$$

$$(4.16) \quad I(1_u) = X(u) \quad \text{a.s.,}$$

$$(4.17) \quad I(g_{n,j}) = \bar{y}_{n,j} \quad \text{a.s.,} \quad (n \in D, j \in S_n)$$

where  $g_{n,j}$  are Haar functions as defined in (2.2) and  $\bar{y}_{n,j}$  as defined in (4.11).

PROOF. (4.14), (4.15) and (4.16) are trivial. To show (4.17) let  $n \in D$  and  $j \in S_n$  be fixed. Since

$$g_{n,j}(u) = (-1)^k 2^{(|n|-p)/2} \quad \text{if } u \in \Delta_{s-\frac{1}{2}, s} \quad \text{and } s \in \bar{V}(n, j, k) \quad \text{for } k = 0, 1, \dots, p$$

$$= 0 \quad \text{otherwise;}$$

and since  $\{\Delta_{s-\frac{1}{2}, s}\} (s \in \bar{V}(n, j))$  are mutually disjoint, we obtain by the definition of the stochastic integral (4.13)

$$I(g_{n,j}) = 2^{(|n|-p)/2} \sum_{k=0}^p (-1)^k \sum_{s \in \bar{V}(n, j, k)} \Delta_{s-\frac{1}{2}, s} X(\omega)$$

$$= \bar{y}_{n,j} \quad \text{a.s.}$$

This completes the proof of the lemma.

From the property (4.15), it follows that

$$(4.18) \quad E[\{I(f)\}^2] = \|f\|^2$$

and this implies that the correspondence between  $f$  and  $I(f)$  is an isometry between  $L^2(A)$  and  $L^2(\Omega)$ . Now suppose that  $f$  is a limit (in  $\| \cdot \|$  norm) of a sequence  $\{f_n\}$  of step functions of the above type. Then l.i.m.  $I(f_n)$  exists defining a random variable  $Y$ . This random variable, as a limit in the mean (in  $\| \cdot \|_{L^2[\Omega]}$  norm), is defined uniquely, neglecting values on a set of zero probability. Also  $Y$  is independent of the particular sequence  $\{f_n\}$  chosen, since two sequences converging to  $f$  in  $\| \cdot \|$

norm can be combined to form a single sequence converging to  $f$  in  $\|\cdot\|$  norm. We define  $\int_A f(u) dX(u)$  as the limit obtained in this way. For  $f \in L^2(A)$   $I(f)$  can be taken to be  $\underline{B}(X)$ -measurable. Now since the family of step functions on  $A$  is dense in  $L^2(A)$ , we have defined a stochastic integral

$$(4.19) \quad I(f) = \int_A f(u) dX(u) \quad \text{for } f \in L^2(A),$$

which satisfies all properties listed in Lemma 3.

We shall show now that the stochastic integral (4.19) can also be defined by a different method. Let  $\{g_{n,j}\}$  ( $n \in D, j \in S_n$ ) be the Haar functions on  $A$  and  $\{\bar{y}_{n,j}\}$  be a sequence of random variables given in (4.11). Consider the series

$$(4.20) \quad \sum_{N=0}^{\infty} \sum_{|n|=N} \sum_{j \in S_n} (f, g_{n,j}) \bar{y}_{n,j} \quad \text{for } f \in L^2(A).$$

The above series converges a.s. by the Three Series Theorem since

$$(4.21) \quad \sum_{N=0}^{\infty} \sum_{|n|=N} \sum_{j \in S_n} E[\{(f, g_{n,j}) \bar{y}_{n,j}\}^2] = \|f\|^2 < \infty.$$

Let us write

$$(4.22) \quad I(f) = \sum_{N=0}^{\infty} \sum_{|n|=N} \sum_{j \in S_n} (f, g_{n,j}) \bar{y}_{n,j} \quad \text{for } f \in L^2(A).$$

We shall accept as  $I(f)$  any random variable equal almost surely to the series on the right.

**THEOREM 4.** *The stochastic integral defined by (4.22) satisfies:*

$$(4.23) \quad I(af + bg) = aI(f) + bI(g)$$

for real numbers  $a$  and  $b$  and  $f, g \in L^2(A)$ .

$$(4.24) \quad E[I(f) \cdot I(g)] = (f, g).$$

$$(4.25) \quad I(f) \text{ is an } N(0, \|f\|^2) \text{ random variable.}$$

$$(4.26) \quad I(g_{n,j}) = \bar{y}_{n,j}$$

where  $g_{n,j}$  are Haar functions on  $A$ .

**PROOF.** (4.23) and (4.26) are obvious from the definition of the stochastic integral. (4.25) follows easily from (4.21) and (4.24) is obtained immediately by using the Parseval's Theorem.

**LEMMA 4.** *Let  $I_1(f)$  and  $I_2(f)$  denote the stochastic integral defined by (4.19) and (4.22) respectively for  $f \in L^2(A)$ . Then*

$$(4.27) \quad I_1(f) = I_2(f) \quad \text{a.s.}$$

**PROOF.** From (4.17) and (4.26),  $I_1(g_{n,j}) = \bar{y}_{n,j} = I_2(g_{n,j})$  a.s. Since  $\{g_{n,j}\}$  is a C.O.N.S. in  $L^2(A)$ , the lemma follows from the isometry given by (4.18) and (4.24).



COROLLARY 2. *Let*

$$(4.28) \quad \bar{X}(u, \omega) = \sum_{N=0}^{\infty} \sum_{|n|=N} \sum_{j \in S_n} \bar{y}_{n,j} G_{n,j}(u) \quad (u \in A)$$

where  $G_{n,j}(u) = (1_u, g_{n,j})$ .

Then  $\bar{X}(u, \omega) = X(u, \omega)$  a.s. for every  $u \in A$ .

PROOF. Since  $\bar{X}(u, \omega) = I_2(1_u)$  a.s. from (4.22) and  $X(u, \omega) = I_1(1_u)$  a.s. from (4.16), the corollary follows immediately from Lemma 4.

The above corollary gives an orthonormal expansion of the  $p$ -parameter Gaussian process. Furthermore, by Theorem 1 this expansion converges uniformly a.s.

We define a random set function

$$(4.29) \quad X(F, \omega) = \int_A 1_F(u) dX(u, \omega) \quad \text{a.s.}$$

for  $F \in \mathcal{A}$ , where  $1_F$  is the indicator function of  $F$ . Then clearly from (4.24)

$$(4.30) \quad E[X(F, \omega)X(F^*, \omega)] = (1_F, 1_{F^*}) = \mu_p(F \cap F^*)$$

for  $F, F^* \in \mathcal{A}$ . Thus  $X(F, \omega)$  is a normal random measure on Lebesgue measure space  $(A, \mathcal{A}, \mu_p)$ . According to Itô [6] it is now possible to define the multiple Wiener integral with respect to the normal random measure  $X(F, \omega)$

$$(4.31) \quad I_q(f_q) = \int_A(q) \int_A f_q(u_1, \dots, u_q) dX(u_1) \cdots dX(u_q)$$

for  $f_q \in L^2(A^q)$  and positive integer  $q \geq 1$ .

**5. The closed linear subspace  $L^*(X(u): u \in A)$ .** Let  $L^*(X(u): u \in A)$  or simply  $L^*(X)$  denote the closed linear subspace in  $L^2(\Omega)$  spanned by all finite linear combinations of the form  $\sum_{i=1}^n c_i X(u_i)$  where  $c_i$ 's are real numbers,  $u_i \in A$ , and  $X(u, \omega)$  is the  $p$ -parameter Gaussian process. Let  $H(R)$  denote the reproducing kernel Hilbert space with the reproducing kernel  $R$ , where  $R$  is the covariance function of the process  $X(u, \omega)$ , with the inner product  $\langle \cdot, \cdot \rangle_{H(R)}$  and norm  $\| \cdot \|_{H(R)}$  (see [1], [13]).

THEOREM 5.

$$(5.1) \quad H(R) \cong_J L^*(X) \quad \text{with} \quad J(R(\cdot, u)) = X(u) \quad (u \in A).$$

Furthermore, if  $f \in H(R)$  and  $J(f) = \xi$ , then

$$(5.2) \quad f(u) = E[\xi \cdot X(u)] \quad \text{for every} \quad u \in A.$$

Here  $H(R) \cong_J L^*(X)$  means that  $J$  is a congruence (inner product preserving isomorphism) from  $H(R)$  onto  $L^*(X)$ .

PROOF. First we note that for each  $f \in H(R)$  there corresponds a unique  $\xi \in L^*(X)$  such that  $f(u) = E[\xi X(u)]$  for  $u \in A$ , since if  $\xi$  and  $\eta$  in  $L^*(X)$  both correspond to  $f$  in  $H(R)$  then  $E[(\xi - \eta)X(u)] = 0$  for every  $u \in A$ , hence  $\xi = \eta$  a.s. as  $X(u)$ , ( $u \in A$ )

span  $L^*(X)$ . Let  $J$  denote such a map, then clearly

$$(5.3) \quad J(R(\cdot, u)) = X(u) \quad \text{for every } u \in A, \quad \text{and}$$

$$(5.4) \quad \langle R(\cdot, u), R(\cdot, v) \rangle_{H(R)} = R(u, v) = E[X(u)X(v)] \quad (u, v \in A).$$

Now  $R(\cdot, u), (u \in A)$  span  $H(R)$  and  $f(u) = \langle f, R(\cdot, u) \rangle_{H(R)} = (\xi, X(u)) = E[\xi X(u)]$ , hence the Basic Congruence Theorem (see [13]) implies (5.1). This completes the proof.

Let  $L^*(1_u: u \in A)$  denote the closed linear space spanned by the elements of the form  $\sum_{i=1}^n c_i 1_{u_i}(\cdot)$  for real numbers  $c_1, \dots, c_n$  and  $u_1, \dots, u_n$  are in  $A$ .

THEOREM 6.

$$(5.5) \quad H(R) \cong_{J_1} L^*(1_u: u \in A) = L^2(A) \quad \text{and}$$

$$J_1(R(\cdot, u)) = 1_u(\cdot) \quad \text{for each } u \in A.$$

Furthermore, if  $f \in H(R)$  and  $J_1(f) = g$ , then

$$(5.6) \quad f(u) = (1_u, g) \quad \text{for each } u \in A.$$

The proof is similar to the proof of Theorem 5.

THEOREM 7.

$$(5.7) \quad L^*(X) = \{I(f): f \in L^2(A)\}.$$

PROOF. Clearly  $I(f) \in L^*(X)$  for each  $f$  in  $L^2(A)$  by the definition of the stochastic integral  $I(f)$ . Let  $\xi \in L^*(X)$ . Then by (5.2) there is an  $f \in H(R)$  corresponding to  $\xi$  with  $f(u) = E[\xi X(u)]$  and in turn there exists  $g$  in  $L^2(A)$  corresponding to  $f$  such that  $f(u) = (g, 1_u)$  by (5.6). Let  $\eta = I(g)$ , then for each  $u \in A$   $f_\eta(u) = E[\eta X(u)] = E[I(g)X(u)] = (g, 1_u) = f(u)$ , i.e.  $E[I(g)X(u)] = E[\xi X(u)]$  for every  $u \in A$ . Hence  $\xi = I(g)$  a.s. and  $\xi \in \{I(f): f \in L^2(A)\}$ .

COROLLARY 3. Let  $\xi$  be in  $L^*(X)$ . Then there exists a function  $g$  in  $L^2(A)$  such that

$$(5.8) \quad \xi = \int_A g(u) dX(u, \omega) \quad \text{a.s.}$$

Furthermore  $g$  can be found under the congruences  $J$  and  $J_1$  given by (5.1) and (5.5) respectively.

**6. Applications.** In this section we shall simply deduce a few results regarding the  $p$ -parameter Gaussian process from the results in [6], [9], [11] and [13].

Let  $(X, B(X))$  be the measurable space where  $X$  is the space of real-valued continuous functions on  $A$  and  $B(X)$  is the  $\sigma$ -field generated by the cylinder sets in  $X$ . Let  $(X, \underline{B}(X), P)$  be the  $p$ -parameter Gaussian process with the mean function zero and the covariance  $R$  where  $\underline{B}(X)$  is the completion of  $B(X)$  under  $P$ .

(A) *A translation theorem.* For  $m \in X$ , the transformation  $\sigma_m: X \rightarrow X$  defined by

$$(6.1) \quad \sigma_m x = x + m$$

clearly sends  $B(X)$ -measurable set into  $B(X)$ -measurable set. The probability measure  $P_m$  given by

$$(6.2) \quad P_m(E) = P(\sigma_m^{-1}E) \quad \text{for } E \in B(X)$$

is Gaussian with the mean function  $m$  and the same covariance function  $R$  as  $P$ . By a direct application of a result from [13] the following theorem is obtained :

**THEOREM 8.**  $P_m \equiv P$  relative to  $B(X)$  if and only if  $m \in H(R)$ . If  $m \in H(R)$ , then the Radon–Nikodym derivative is given by

$$(6.3) \quad \frac{dP_m}{dP}(x) = \exp \left\{ u_m(x) - \frac{1}{2} \|m\|_{H(R)}^2 \right\}$$

where  $u_m(x)$  is in  $L^*(X)$  which corresponds to  $m \in H(R)$  under the congruence of (5.1).

The notation  $P_m \equiv P$  means that  $P_m$  and  $P$  are mutually absolutely continuous.

Now from Corollary 3 and (5.6) there exists  $g$  in  $L^2(A)$  which corresponds to  $m$  in  $H(R)$  such that  $u_m(x) = \int_A g(u) dx(u)$  and  $\|m\|_{H(R)} = \|g\|$ . Hence the Radon–Nikodym derivative in (6.3) becomes

$$(6.4) \quad \frac{dP_m}{dP}(x) = \exp \left\{ \int_A g(u) dx(u) - \frac{1}{2} \|g\|^2 \right\}.$$

For  $p = 2$  similar results are obtained in Park [12] and Yeh [16] but their approaches differ from ours.

**(B) Equivalence of Gaussian measures.** Let  $(X, B(X), Q)$  be a Gaussian measure space with the mean function  $m$  and the covariance function  $\Gamma_Q$ . Then we can deduce the following theorem from a result of Oodaira [11] using the same method as Kailath [7].

**THEOREM 9.**  $Q \equiv P$  relative to  $B(X)$  if and only if there is a symmetric kernel  $K \in L^2(A \times A)$  such that

$$(6.5) \quad \Gamma_Q(u, v) = R(u, v) - \int_{A \times A} 1_u(s) 1_v(t) K(s, t) d\mu_p(s) d\mu_p(t),$$

$$(6.6) \quad i \notin \sigma(K) \quad \text{and}$$

$$(6.7) \quad m \in H(R)$$

where  $\sigma(K)$  denotes the spectrum of the operator  $K$  given by

$$(6.8) \quad (Kf)(u) = \int_A K(s, u) f(s) d\mu_p(s), \quad u \in A \quad \text{and } f \in L^2(A).$$

We note that if  $m \in H(R)$  there is a function  $k$  in  $L^2(A)$  such that  $m(u) = \int_A 1_u(s) k(s) d\mu_p(s)$ , by Theorem 6. Let  $\lambda_j$  and  $\varphi_j$  ( $j = 1, 2, \dots$ ) be eigenvalues and eigenfunctions respectively of the operator  $K$  and let  $D(\cdot)$  be the Fredholm determinant of  $K$ , i.e.

$$(6.9) \quad D(\lambda) = \prod_{j=1}^{\infty} (1 - \lambda \lambda_j).$$

For each value of  $\lambda$  for which  $\lambda^{-1} \notin \sigma(K)$  there exists a unique kernel  $H_\lambda \in L^2(A \times A)$  called the Fredholm resolvent of  $K$  at  $\lambda$  satisfying

$$(6.10) \quad H_\lambda - K = \lambda H_\lambda K = \lambda K H_\lambda.$$

We denote  $H_1$  by  $H$ .

THEOREM 10. *If  $Q \equiv P$ , then*

$$(6.11) \quad \frac{dQ}{dP}(x) = D(1)^{-\frac{1}{2}} \exp \left\{ -\frac{1}{2} I_2(H) - \frac{1}{2} \int_A H(s, s) d\mu_p(s) - \frac{1}{2} \|k\| - \frac{1}{2} (Hk, k) + I(k) + \int_{A \times A} H(s, t) k(s) d\mu_p(s) dx(t) \right\}$$

where  $I_2(\cdot)$  is the 2nd degree multiple Wiener integral.

PROOF. Define

$$(6.12) \quad \zeta_j = \int_A \varphi_j(s) dx(s).$$

Then clearly  $\zeta_j$  are independent  $N(0, 1)$  random variables under  $P$  and independent  $N(k_j, 1 - \lambda_j)$  random variables under  $Q$  where  $k_j = (k, \varphi_j)$ .

Let

$$(6.13) \quad F_j(x) = (1 - \lambda_j)^{-\frac{1}{2}} \exp \left[ -\frac{1}{2} (\zeta_j - k_j)^2 / (1 - \lambda_j) \right] / \exp \left[ -\frac{1}{2} \zeta_j^2 \right],$$

then it can be shown by using the same method as Shepp [14]:

$$(6.14) \quad F(x) = \prod_{j=1}^{\infty} F_j(x) \text{ converges a.s. } (P),$$

$$(6.15) \quad F(x) = \frac{dQ}{dP}(x),$$

$$(6.16) \quad \sum_{j=1}^{\infty} k_j^2 / (1 - \lambda_j) = (k, k) + (Hk, k),$$

$$(6.17) \quad \sum_{j=1}^{\infty} \lambda_j \zeta_j^2 / (1 - \lambda_j) = I_2(H) + \int_A H(s, s) d\mu_p(s), \quad \text{and}$$

$$(6.18) \quad \sum_{j=1}^{\infty} k_j \zeta_j / (1 - \lambda_j) = I(k) + \int_{A \times A} H(s, t) k(s) d\mu_p(s) dx(t).$$

Hence we can obtain (6.11) and the theorem is proved.

Shepp [14] has recently proved the same results for the standard Wiener process.

(C) *A zero-one law.* We shall obtain a zero-one law for the  $r$ -module of the  $p$ -parameter Gaussian process by a direct application of the result of Kallianpur [9], which we shall state here first.

Let  $Q$  be a Gaussian probability measure on the measurable space  $(X, B(X))$ , where  $X$  is a family of real-valued functions  $x(\cdot)$  defined on a set  $T$ ,  $B(X)$  is the  $\sigma$ -field generated by the cylinder sets in  $X$ , and  $\underline{B}(X)$  is its completion under  $Q$ . Let  $\Gamma_Q$  denote the covariance function of  $Q$  and assume the mean function to be zero.

A subset  $M$  of  $X$  is called an  $r$ -module if for every  $x_1$  and  $x_2$  in  $M$  and rational numbers  $r_1$  and  $r_2$ ,  $r_1x_1 + r_2x_2 \in M$ . Kallianpur [9] shows that if  $M$  is a  $\underline{B}(X)$ -measurable  $r$ -module, then  $Q(M) = 0$  or  $1$ , under the following general assumptions:

(6.19)  $T$  is a complete separable metric space,

(6.20)  $X$  is a linear space of functions under the usual operation of addition of functions and multiplication by real scalars,

(6.21)  $\Gamma_Q$  is continuous on  $T \times T$ ,

(6.22)  $H(\Gamma_Q) \subset X$ .

**THEOREM 11.** *Let  $(X, \underline{B}(X), P)$  be the  $p$ -parameter Gaussian process and let  $M$  be a  $\underline{B}(X)$ -measurable  $r$ -module. Then  $P(M) = 0$  or  $1$ .*

**PROOF.** The assumptions (6.19)–(6.21) are clearly satisfied by the process  $(X, \underline{B}(X), P)$ . Let  $h \in H(R)$ . Then there exists a function  $g \in L^2(A)$  such that  $h(u) = (1_u, g)$  by (5.6). Now  $h$  is clearly continuous and  $h(u) = 0$  for  $u \in A_0$ , i.e.,  $h \in X$ . Thus (6.22) is satisfied and the conclusion follows from [9].

(D) *Homogeneous chaos.* We shall give an orthogonal expansion of the  $L^2$ -functional of the  $p$ -parameter Gaussian process  $(X, \underline{B}(X), P)$ . It is easy to see that  $\xi$  is a  $L^2$ -functional of the normal random measure  $X(F, \omega)$  (in the sense of Itô [6]) if and only if  $\xi \in L^2(X, \underline{B}(X), P)$ . The following theorem is deduced from Itô [6].

**THEOREM 12.** *Let  $\xi \in L^2(X, \underline{B}(X), P)$ . Then  $\xi$  can be expressed in the form:*

$$(6.23) \quad \xi = \sum_{q \geq 0} I_q(f_q)$$

where  $I_q(\cdot)$  is the multiple Wiener integral,  $f_q$  is given by the following orthogonal development

$$(6.24) \quad f_q(u_1, \dots, u_q) = 2^{\frac{1}{2}} \sum_{n_1 + \dots + n_r = q} \sum_{\lambda_1, \dots, \lambda_r} a_{\lambda_1}^{n_1} \dots a_{\lambda_r}^{n_r} \\ g_{\lambda_1}(u_1) \dots g_{\lambda_1}(u_{n_1}) g_{\lambda_2}(u_{n_1+1}) \dots g_{\lambda_2}(u_{n_1+n_2}) \dots \\ g_{\lambda_r}(u_{n_1+\dots+n_{r-1}+1}) \dots g_{\lambda_r}(u_{n_1+\dots+n_r}),$$

$\{g_{\lambda_j}\}$  is a C.O.N.S. in  $L^2(A)$ , and  $\{a_{\lambda_j}^{n_1} \dots a_{\lambda_r}^{n_r}\}$  is the Fourier Hermite coefficient given in [2].

#### REFERENCES

- [1] ARONSZAJN, N. (1950). Theory of reproducing kernels. *Amer. Math. Soc. Trans.* **68** 337–404.
- [2] CAMERON, R. H. and MARTIN, W. T. (1947). The orthogonal development of non-linear functionals in series of Fourier Hermite functions. *Ann. of Math.* 385–392.
- [3] CIESIELSKI, Z. (1961). Holder condition for realization of Gaussian processes. *Amer. Math. Soc. Trans.* **99** 403–413.
- [4] DOOB, J. L. (1953). *Stochastic Processes*. Wiley, New York.
- [5] HAAR, A. (1910). Zur theorie der orthogonalen funktionen-systeme. *Math. Ann.* **69** 331–371.
- [6] ITÔ, K. (1951). Multiple Wiener integral. *J. Math. Soc. Japan* **3** 157–169.

- [7] KAILATH, T. (1967). On measures equivalent to Wiener measure. *Ann. Math. Statist.* **38** 261–263.
- [8] KAKUTANI, S. (1948). On equivalence of infinite product measures. *Ann. of Math.* **49** 214–224.
- [9] KALLIANPUR, G. (1970). Zero-one law for Gaussian processes. *Amer. Math. Soc. Trans.* **149**.
- [10] KUELBS, J. D. (1965). Integration on space of continuous functions. Ph.D. thesis, Univ. of Minnesota.
- [11] OODAIRA, H. (1963). The equivalence of Gaussian stochastic processes. Ph.D. thesis, Michigan State Univ.
- [12] PARK, CHULL (1968). Generalized Riemann–Stieltjes integral over K.Y.W. space of functions of two variables. Ph.D. thesis, Univ. of Minnesota.
- [13] PARZEN, E. (1959). Statistical inference on time series by Hilbert space methods, I. Technical Report No. 23, Department of Statistics, Stanford Univ.
- [14] SHEPP, L. A. (1960). Radon–Nikodym derivative of Gaussian measures. *Ann. Math. Statist.* **37** 321–354.
- [15] YEH, J. (1960). Wiener measure in a space of functions of two variables. *Amer. Math. Soc. Trans.* **95** 433–450.
- [16] YEH, J. (1963). Cameron–Martin translation theorems in the Wiener space of function of two variables. *Amer. Math. Soc. Trans.* **107** 409–420.
- [17] YEH, J. (1963). Orthogonal development of functional and related theorems in the Wiener space of functions of two variables. *Pacific J. Math.* **13** 1427–1436.
- [18] ZAAANEN, A. C. (1964). *Linear Analysis*. North-Holland, Amsterdam.