

## ON THE INVARIANCE PRINCIPLE AND FUNCTIONAL LAW OF ITERATED LOGARITHM FOR NONERGODIC PROCESSES

By

DALIBOR VOLNÝ

(Received April 30, 1987)

**Abstract:** Recently, nonergodic versions of several limit theorems for strictly stationary processes were given ([6], [7], [8]). The proofs were based on the fact that important properties of the stationary process which hold with respect to an invariant measure  $\mu$ , are preserved with respect to its ergodic components. Using the same technique we shall prove two nonergodic versions of Heyde's functional CLT and log log law here.

### 1. Basic notions and the main results.

Let  $(\Omega, \mathcal{A}, T, \mu)$  be a dynamical system where  $(\Omega, \mathcal{A}, \mu)$  is a probability space with  $\sigma$ -algebra  $\mathcal{A}$  of subsets of  $\Omega$ ,  $\mu$  is a probability measure on the measure space  $(\Omega, \mathcal{A})$  and the mapping  $T: \Omega \rightarrow \Omega$  is bijective and bimeasurable,  $\mu T^{-1} = \mu$ . The collection  $\mathcal{I}$  of all sets  $A \in \mathcal{A}$  such that  $A = TA$  is a  $\sigma$ -algebra. If  $\mu(A) = 0$  or  $\mu(A) = 1$  for each  $A \in \mathcal{I}$ , we say that  $\mu$  is ergodic. A  $\sigma$ -algebra  $\mathcal{M} \subset \mathcal{A}$  for which  $\mathcal{M} \subset T^{-1}\mathcal{M}$  is called invariant (sometimes, an invariant  $\sigma$ -algebra is defined by the property  $T^{-1}\mathcal{M} \subset \mathcal{M}$ ). Let us denote  $\mathcal{M}_i = T^{-i}\mathcal{M}$ ,  $i \in \mathbb{Z}$ ;  $\mathcal{M}_{-\infty} = \bigcap_{i \in \mathbb{Z}} \mathcal{M}_i$  and  $\mathcal{M}_{\infty}$  is the coarsest  $\sigma$ -algebra containing all  $\mathcal{M}_i$ ,  $i \in \mathbb{Z}$ .

For a measurable function  $f$ ,  $(f \circ T^i)$  is a strictly stationary process. Let us suppose that  $f$  is square integrable and let us denote  $X_i = f \circ T^i$ ,  $S_n = \sum_{i=1}^n X_i$ ,  $\sigma_n^2 = E(S_n^2 | \mathcal{I})$ . Let  $\varphi(t) = (2t \log \log t)^{1/2}$ ,  $e < t < \infty$ . For  $\sigma_n^2 > 0$  a.s. we define

$$\theta_n(t) = \sigma_n^{-1}(S_k + (nt - k)X_{k+1}), \quad k \leq nt \leq k+1, \quad k=0, 1, \dots, n-1;$$

$$\eta_n(t) = (\varphi(\sigma_n^2))^{-1}(S_k + (nt - k)X_{k+1}) \quad \text{if } \sigma_n^2 > e,$$

$$= 0 \quad \text{otherwise, } k \leq nt \leq k+1, \quad k=0, 1, \dots, n-1.$$

We put  $g = \sup\{n : \sigma_n^2 \leq e\}$ . Notice that for  $\mu$  ergodic, the functions  $\sigma_n$  and  $g$  are constant almost surely; in Theorem 1 we shall consider them as numbers.

The functions  $\theta_n$  and  $\eta_n$  belong to  $C[0, 1]$ , the space of all continuous functions on  $[0, 1]$ . Let  $K$  be the set of all absolutely continuous functions

$x \in C[0, 1]$  such that  $x(0) = 0$  and  $\int_0^1 [\dot{x}(t)]^2 dt \leq 1$  and  $\dot{x}$  denotes the derivative of  $x$  determined almost everywhere with respect to the Lebesgue measure. By  $W$  we denote the standard Brownian motion on  $[0, 1]$  and  $\xrightarrow{d}$  denote the convergence of  $C[0, 1]$ -valued r. v. in distribution as defined in [1].

For  $p \in \mathbb{Z}$ , let  $x_p = E(X_{-p} | \mathcal{M}_0) - E(X_{-p} | \mathcal{M}_{-p})$ .

**Theorem 1.** (C. C. Heyde). *Let the measure  $\mu$  be ergodic and*

- (1)  $E(X_0 | \mathcal{M}_\infty) = X_0$ ,  $E(X_0 | \mathcal{M}_{-\infty}) = 0$ ,
- (2)  $\sum_{m=1}^{\infty} \{ \limsup_{n \rightarrow \infty} E(\sum_{p=m}^n x_p)^2 + \limsup_{n \rightarrow \infty} E(\sum_{p=-m}^n x_{-p})^2 \} < \infty$ .

*Then there exists a limit  $\sigma = \lim_{n \rightarrow \infty} \sigma_n / n^{1/2}$  and  $0 \leq \sigma < \infty$ . If  $\sigma > 0$  then  $\theta_n \xrightarrow{d} W$ . Also,  $g < \infty$ ,  $\{\eta_n : n > g\}$  is relatively compact, and the set of its limit points coincides with  $K$ .*

The proof is given in [2], [3].

Here we shall prove nonergodic version of Theorem 1. In that case,  $\sigma_n$  and  $\sigma$  should be considered as  $\mathcal{G}$ -measurable functions.

**Theorem 2.** *Let the assumptions (1) and (2) of Theorem 1 hold. Then there exists a limit  $\sigma = \lim_{n \rightarrow \infty} \sigma_n / n^{1/2}$  and  $0 \leq \sigma < \infty$ . If  $\sigma > 0$  a. s. then  $\theta_n \xrightarrow{d} W$ , the family  $\{\eta_n ; n \in \mathbb{N}\}$  is relatively compact a. s., and the set of its limit points coincides with  $K$ .*

**Theorem 3.** *Let the assumption (1) of Theorem 1 hold and let*

- (3)  $\sum_{m=1}^{\infty} \{ \limsup_{n \rightarrow \infty} E((\sum_{p=m}^n x_p)^2 | \mathcal{G}) + \limsup_{n \rightarrow \infty} E((\sum_{p=-m}^n x_{-p})^2 | \mathcal{G}) \} < \infty$ .

*Then the conclusion of Theorem 2 holds.*

## 2. Some auxiliary results and the proofs of Theorems 2, 3.

Let the invariant  $\sigma$ -algebra  $\mathcal{M}$  be given and let the square integrable function  $f$  satisfy the assumptions of Theorem 2 or 3. We shall show that without loss of generality we can suppose that the dynamical system  $(\Omega, \mathcal{A}, T, \mu)$  has some useful properties, namely that there exists an ergodic decomposition of the measure  $\mu$ . The proofs of results presented here can be found in [5], [6], [7], [8].

We say that a  $\sigma$ -algebra  $\mathcal{E} \subset \mathcal{A}$  is separable if there exists an at most countable collection of sets generating  $\mathcal{E}$  (see [4]). It can be easily shown that there exists a separable  $\sigma$ -algebra  $\mathcal{M}^* \subset \mathcal{M}$  such that  $E(X_i | T^j \mathcal{M}^*) = E(X_i | T^j \mathcal{M})$  a. s. for all  $i, j \in \mathbb{Z}$ . Hence, there exists a separable  $\sigma$ -algebra  $\mathcal{E} \subset \mathcal{A}$  such that

$T^{-1}\mathcal{E}=\mathcal{E}$ ,  $\mathcal{M}_i^* \subset \mathcal{E}$  and  $X_i, E(X_i|T^j\mathcal{M})$  are  $\mathcal{E}$ -measurable,  $i, j \in \mathbb{Z}$ . We can thus suppose that  $\mathcal{A}$  and  $\mathcal{M}$  are separable. Then, there exists a function  $g$  such that  $\mathcal{A}$  is generated by  $\{g^{-1}(A): A \subset \mathbb{R}$  is an interval $\}$  (we can take  $g = \sum_{j=1}^{\infty} \frac{1}{2^j} \chi_{A_j}$  where  $\{A_1, A_2, \dots\}$  generates  $\mathcal{A}$ ).

Let us define  $\phi: \Omega \rightarrow \mathbb{R}^{\mathbb{Z}}$  by  $(\phi\omega)_i = g(T^i\omega)$ ,  $i \in \mathbb{Z}$ . Let  $S$  be a shift in  $\mathbb{R}^{\mathbb{Z}}$ , i. e.  $(Sz)_i = z_{i+1}$ . We have  $\phi \circ T = S \circ \phi$  and  $\phi^{-1}(\mathcal{B}^{\mathbb{Z}}) = \mathcal{A}$ ,  $\mathcal{B}^{\mathbb{Z}}$  being the Borel  $\sigma$ -algebra in  $\mathbb{R}^{\mathbb{Z}}$ . The probability measure  $\nu = \mu\phi^{-1}$  is  $S$ -invariant, i. e.  $\nu = \nu S^{-1}$ . So,  $(\mathbb{R}^{\mathbb{Z}}, \mathcal{B}^{\mathbb{Z}}, S, \nu)$  is a dynamical system. For an  $\mathcal{A}$ -measurable function  $h$  on  $\Omega$  there exists a  $\mathcal{B}^{\mathbb{Z}}$ -measurable function  $\bar{h}$  on  $\mathbb{R}^{\mathbb{Z}}$ ,  $h = \bar{h} \circ \phi$ . If  $h$  is integrable and  $\bar{\mathcal{E}} \subset \mathcal{B}^{\mathbb{Z}}$  is a  $\sigma$ -algebra, it is  $E_{\mu}(h|\phi^{-1}\bar{\mathcal{E}}) = E_{\nu}(\bar{h}|\bar{\mathcal{E}}) \circ \phi$  a. s. ( $\mu$ ). For the  $\sigma$ -algebra  $\bar{\mathcal{I}} = \{A \in \mathcal{B}^{\mathbb{Z}}: A = SA\}$  we have  $\mathcal{I} = \phi^{-1}(\bar{\mathcal{I}})$ . So, we can assume that  $(\Omega, \mathcal{A}, T, \mu) = (\mathbb{R}^{\mathbb{Z}}, \mathcal{B}^{\mathbb{Z}}, S, \nu)$ . Under this assumption there exists a family  $(m_{\omega}; \omega \in \Omega)$  of regular conditional probabilities induced by  $\mathcal{I}$  with respect to  $\mu$ ; moreover,  $m_{\omega} T^{-1} = m_{\omega}$  and the measures  $m_{\omega}$  are ergodic for almost all ( $\mu$ )  $\omega \in \Omega$  (see [5]).

Let  $\mathcal{E} \subset \mathcal{A}$  be a  $\sigma$ -algebra and let  $\nu$  be a probability measure on the measure space  $(\Omega, \mathcal{A})$ .  $L^2(\mathcal{E}, \nu)$  denotes the Hilbert space of all  $\mathcal{A}$ -measurable functions  $g$  such that  $\int g^2 d\nu < \infty$  and there exists a  $\mathcal{E}$ -measurable function  $h$ ,  $g = h$  a. s. ( $\nu$ ). Functions from  $L^2(\mathcal{E}, \nu)$  which are equal almost surely ( $\nu$ ) are considered as equal.

The projection operator onto  $L^2(\mathcal{M}_i, \mu) \ominus L^2(\mathcal{M}_{i-1}, \mu)$  is denoted by  $P_i$ ,  $i \in \mathbb{Z}$ . We have  $P_i h = E(h|\mathcal{M}_i) - E(h|\mathcal{M}_{i-1})$ ,  $h \in L^2(\mathcal{A}, \mu)$ ; so,  $x_p = P_0 X_{-p}$ . By  $P_i^{\omega}$  we denote the projection operator onto  $L^2(\mathcal{M}_i, m_{\omega}) \ominus L^2(\mathcal{M}_{i-1}, m_{\omega})$ .

**Lemma 1.** *Let  $h \in L^2(\mathcal{M}_{\infty}, \mu)$ . Then for almost all ( $\mu$ )  $\omega \in \Omega$  we have  $h \in L^2(\mathcal{M}_{\infty}, m_{\omega})$  and  $E(h|\mathcal{M}_i) = E_{m_{\omega}}(h|\mathcal{M}_i)$  a. s. ( $m_{\omega}$ ),  $E(h^2|\mathcal{M}_i) = E_{m_{\omega}}(h^2|\mathcal{M}_i)$  a. s. ( $m_{\omega}$ ),  $P_i^{\omega} h = P_i h$  a. s. ( $m_{\omega}$ ).*

The proof of Lemma 1 is given in [5].

**Corollary.** *Let  $h \in L^2(\mathcal{A}, \mu)$  and  $h, h \circ T, h \circ T^2, \dots$  be a martingale difference sequence in  $L^2(\mathcal{A}, \mu)$ . Then for almost all ( $\mu$ )  $\omega \in \Omega$ ,  $h, h \circ T, h \circ T^2, \dots$  is a martingale difference sequence in  $L^2(\mathcal{A}, m_{\omega})$ .*

**Proof.** The  $\sigma$ -algebra generated by  $h \circ T^{-n}$ ,  $n \geq 0$ , is separable and invariant. We can thus assume that  $h = P_0 h$ . Following Lemma 1 we have  $h = P_0^{\omega} h$  a. s. ( $m_{\omega}$ ), hence  $(h \circ T^i)$  is a martingale difference sequence in  $L^2(\mathcal{A}, m_{\omega})$  for a. e. ( $\mu$ )  $\omega \in \Omega$ .

Let  $h_n$ ,  $n = 1, 2, \dots$  be  $C[0, 1]$ -valued random variables. We can view them as functions  $h_n(t, \omega)$  of  $t$  and  $\omega$ . The variables  $h_n$  converge to  $W$  in distribution

iff their finite-dimensional distributions converge to those of  $W$  and the sequence  $(h_n)$  is tight. The first condition can be expressed by the use of characteristic functions:

(i) For each finite  $k$ , each collection  $0 \leq t_1 < \dots < t_k \leq 1$ , and  $(s_1, \dots, s_k) \in \mathbf{R}^k$  it is

$$\int \exp(i \sum_{j=1}^k s_j \cdot h_n(t_j, \cdot)) d\mu \xrightarrow{n \rightarrow \infty} \exp\left(-\frac{1}{2}(t_1 \cdot (\sum_{j=1}^k s_j)^2 + (t_2 - t_1) \cdot (\sum_{j=2}^k s_j)^2 + \dots)\right).$$

The second condition can be expressed by (ii), (iii), compare [2], as follows:

(ii)  $\sup_n \mu(|h_n(0, \cdot)| > \lambda) \xrightarrow{\lambda \rightarrow \infty} 0$ .

(iii) For each  $\varepsilon > 0$ ,  $\sup_n \mu(\sup_{|s-t| < \delta} |h_n(s, \cdot) - h_n(t, \cdot)| > \varepsilon) \xrightarrow{\delta \downarrow 0} 0$ .

The convergence of  $h_n$  is related to the measure  $\mu$ ; in the case described above we say that  $h_n$  converge in distribution with respect to  $\mu$ .

**Lemma 2.** *Let for almost all  $(\mu)$   $\omega \in \Omega$ ,  $h_n$  converge to  $W$  in distribution with respect to  $m_\omega$ . Then  $h_n$  converge to  $W$  in distribution with respect to  $\mu$ .*

**Proof.** Condition (i) immediately follows from the fact that

$$\int \exp(i \cdot \sum_{j=1}^k s_j \cdot h_n(t_j, \cdot)) d\mu = \iint \exp(i \cdot \sum_{j=1}^k s_j \cdot h_n(t_j, \cdot)) dm_\omega d\mu(\omega).$$

Let  $\varepsilon > 0$  be fixed. For  $\tau > 0$  we put

$$A(\lambda, \tau) = \mu\{\omega : \sup_n m_\omega(|h_n(0, \cdot)| > \lambda) < \tau\}$$

and

$$B(\delta, \tau) = \mu\{\omega : \sup_n m_\omega(\sup_{|s-t| < \delta} |h_n(s, \cdot) - h_n(t, \cdot)| > \varepsilon) < \tau\}.$$

For each  $n$  it holds  $\mu(|h_n(0, \cdot)| > \lambda) \leq \tau + 1 - A(\lambda, \tau)$  and

$$\mu(\sup_{|s-t| < \delta} |h_n(s, \cdot) - h_n(t, \cdot)| > \varepsilon) \leq \tau + 1 - B(\delta, \tau).$$

We have  $A(\lambda, \tau) \xrightarrow{\lambda \rightarrow \infty} 1$  and  $B(\delta, \tau) \xrightarrow{\delta \downarrow 0} 1$  hence the conditions (ii), (iii) hold for  $\mu$ , too.

**Proof of Theorem 2.** Following [2], pp. 141-142, there exist functions  $g, h \in L^2(\mathcal{A}, \mu)$  such that  $f = g + h - h \circ T$  and  $(g \circ T^i)$  is a martingale difference sequence. In the ergodic case, the theorem can be derived from the existence of this decomposition (see [3], [2]). According to the Corollary to Lemma 1, the decomposition exists in a.e.  $(\mu)$  probability space  $(\Omega, \mathcal{A}, m_\omega)$ . So, the functional CLT and log log law hold in a.e.  $(\mu)$  probability space  $(\Omega, \mathcal{A}, m_\omega)$ . We get the invariance principle in  $(\Omega, \mathcal{A}, \mu)$  by Lemma 2. The proof of the functional log log law is even easier: Let  $A$  be the set of all elements of  $\Omega$

for which  $\{\eta_n : n \in \mathbb{N}\}$  is relatively compact and has  $K$  as the set of limit points. We have  $m_\omega(A) = 1$  for a.e.  $(\mu)$   $\omega \in \Omega$ , so  $\mu(A) = 1$  which finishes the proof of Theorem 2.

**Proof of Theorem 3.** According to Lemma 1, the assumptions of Theorem 1 are fulfilled in almost all  $(\mu)$  probability spaces  $(\Omega, \mathcal{A}, m_\omega)$ . Similarly as in the proof of Theorem 2 we get the desired result from its validity in spaces  $(\Omega, \mathcal{A}, m_\omega)$ .

### References

- [1] Billingsley, P.: *Convergence of Probability Measures*. Wiley, New York, (1968).
- [2] Hall, P. and Heyde, C.C.: *Martingale Limit Theory and its Application*. Academic Press, New York, (1980).
- [3] Heyde, C.C.: On the central limit theorem and iterated logarithm law for stationary processes. *Bull. Austral. Math. Soc.* 12 (1975), 1-8.
- [4] Loève, M.: *Probability Theory*. Van Nostrand, Princeton, (1960).
- [5] Volný, D.: Martingale decompositions of stationary processes. *Yokohama Math. J.* 35 (1987), 113-121.
- [6] Volný, D.: The central limit problem for strictly stationary sequences. Ph.D. Thesis, Math. Institute of Charles Univ., Prague (1984).
- [7] Volný, D.: A non-ergodic version of Gordin's CLT for integrable stationary processes. *Comm. Math. Univ. Carolinae* 28,3 (1987), 413-419
- [8] Volný, D. and Yokoyama, R.: On the law of iterated logarithm for martingales. Submitted for publication.

The Mathematical Institute of  
Charles University  
Sokolovská 83, 18600  
Praha 8, Czecholovakia