

# Spectral Representation of Multiply Self-decomposable Stochastic Processes\*

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Abstract. In the present paper we prove an integral representation for multiply self-decomposable processes which is similar to a known representation of Rajput, B.S. and Rosinski J. [3] for Gaussian, stable and infinitely divisible processes.

## I. Introduction, notation and preliminaries.

The classical spectral representation theory for Gaussian processes has been widely studied and applied in

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various branches of sciences such as prediction and filtration problems, signal transmissions, biological models, quantum mechanics, ... It is then a natural idea to develop the theory for wider classes of processes containing stable processes, semi-stable processes and infinitely divisible (i.d.) processes. In the sequel, we study the above problem for multiply self-decomposable (shortly, m.s.d.) real-valued processes only although our method as well as results are true for general separable Banach spaces.

We say that a stochastic process  $X_t, t \geq 0$ , is a Lévy process (cf. Sato [4]), if

- (i)  $X(0) = 0$ , (P.1);
- (ii) It is an independent increment process;
- (iii) It is temporally homogeneous;
- (iv) With P.1 its realizations are CADLAG, i.e. continuous from the right having the left limits.

Let  $\mathbb{P}$  denote the class of all p.m.'s on the  $\sigma$ -field  $\mathcal{B}$  of Borel subsets of the real line  $\mathbb{R}$  equipped with the weak convergence. Given a positive number  $c$  and a p.m.  $\mu \in \mathbb{P}$  let  $T_c\mu$  denote a p.m. given by

$$(1) \quad (T_c\mu)(E) = \mu(\{c^{-1}x : x \in E\}),$$

where  $E \in \mathcal{B}$ .

Suppose that  $X_1, X_2, \dots$  is a sequence of real-valued independent r.v.'s and  $\{a_n\}, \{b_n\}$  are sequences of real

numbers such that  $a_n > 0, n = 1, 2, \dots$  and the triangular array  $X_{n,k} := a_n X_k$  obey the uniformly asymptotically negligible (UAN), that is  $X_{n,k} \rightarrow 0$  uniformly in  $k$  in probability. The class  $L = L_1$  of possible shift limit laws  $\mu$  (distributions) of the sum  $S_n = \sum_{k=1}^n X_{n,k}$  is non-empty and consists of all *self-decomposable p.m.'s on  $\mathbb{R}$*  (cf. Loe've [2], p. 323).

Recall that a p.m.  $\mu \in \mathbb{P}$  is called *self-decomposable (shortly, s.d.)*, if for every  $c \in (0, 1)$  there exists a p.m.  $\mu_c$  such that

$$(2) \quad \mu = (T_c \mu) * \mu_c,$$

where  $*$  denotes the ordinary convolution of p.m.'s.

It is known ([2], p. 323) that if  $\mu$  is s.d. then  $\mu$  and  $\mu_c$  are both i.d. Let  $L_0$  denote the class of all i.d.p.m.'s on  $\mathbb{R}$ . The classes  $L_n, n = 1, 2, \dots$  of  $n$ -times s.d.p.m.'s were first introduced and studied by Urbanik [7] and then studied further by many other authors (cf., for example, [1,5,6]). They are defined recursively as follows: A p.m.  $\mu \in L_n, n = 2, 3, \dots$  if and only if  $\mu \in L_1$  and for each  $c \in (0, 1)$  the component  $\mu_c$  in (1.2) belongs to  $L_{n-1}$ .

It has been proved by Thu ([6], Proposition 1.1) that a p.m.  $\mu \in L_n, n = 1, 2, \dots$ , if and only if, for every

$c \in (0, 1)$  there exists a p.m.  $\nu := \mu_{c,n} \in L_0$  such that the following equality holds:

$$(3) \quad \mu = *_{k=1}^{\infty} (T_{c^k} \nu)^{*r_{k,n}},$$

where the power is taken in the convolution sense and, for  $n=1,2, \dots$ ;  $k=0,1,2,\dots$  we put

$$(4) \quad r_{k,n} = \binom{n+k-1}{k}.$$

The formulas (3) and (4) lead to the following interpolation of classes  $L_n$  ( cf.Thu [6]):

For each  $\alpha > 0$  we put

$$(5) \quad \binom{\alpha}{k} = \begin{cases} 1 & k = 0, \\ \alpha(\alpha-1)\dots(\alpha-k+1)/k! & k = 1, 2, \dots \end{cases}$$

and introduce the class  $\alpha$ -times s.d. p.m.'s as the following:

**1. Definition** (cf.Thu [6]). *A p.m.  $\mu \in L_\alpha, \alpha > 0$ , if and only if, for every  $c \in (0, 1)$  there exists a p.m.  $\nu := \mu_{c,\alpha} \in L_0$  such that the following equality holds:*

$$(6) \quad \mu = *_{k=1}^{\infty} (T_{c^k} \nu)^{*r_{k,\alpha}},$$

where, the power is taken in the convolution sense and, for  $\alpha > 0; k = 0, 1, 2, \dots$  we put

$$(7) \quad r_{k,\alpha} = \binom{\alpha+k-1}{k}.$$

It should be noted that the infinite convolution (6) is weakly convergent if and only if

$$(8) \quad \int_{-\infty}^{\infty} \log^{\alpha}(1 + |x|)\nu(dx) < \infty.$$

The above definition is equivalent to the following two definitions:

**1. Definition** (cf. Thu [6]). A p.m.  $\mu \in L_{\alpha}, \alpha > 0$  if and only if, for every  $c \in (0, 1)$  there exists a p.m.  $\nu := \mu_{c, \alpha} \in L_0$  such that the condition (5) is satisfied and  $\mu$  is the distribution of the following random series

$$(9) \quad \sum_{k=0}^{\infty} Z_k c^k,$$

where  $Z_k, k = 0, 2, 3, \dots$  are independent r.v.'s with distributions  $\nu^{*r_{k, \alpha}}$ , respectively.

**1''.** **Definition** (cf. Hong [1]). A p.m.  $\mu \in L_{\alpha}, \alpha > 0$  if and only if there exists a Lévy process  $(X(t))$  such that

$$(10) \quad \mu \stackrel{d}{=} \int_0^{\infty} \exp(t^{1/\alpha})X(dt),$$

where the Lévy process  $(X(t))$  must satisfy the condition

$$E \log^{\alpha}(1 + |X(1)|) < \infty.$$

It should be noted that for any  $0 < \alpha < \beta$  we have  $L_{\beta} \subset L_{\alpha}$  and the intersection  $L_{\infty}$  of all classes  $L_{\alpha}, \alpha >$

0 is non-void, since it contains Gaussian and stable p.m.'s. The p.s.'s in the class  $L_\infty$  are called completely s.d., or, mixed stable ( cf. Thu [5,6], Urbanik [7]).

In the sequel we shall need the following representation of i.d. and m.s.d.p.m.'s:

**2. Theorem** (cf. Loéve [2]) A p.m.  $\mu$  is i.d. if and only if its characteristic function  $\hat{\mu}(t)$  is of the unique form:

$$(11) \quad \begin{cases} -\log \hat{\mu}(t) = iat + \sigma^2 t^2 \\ -\int_{-\infty}^{\infty} (e^{itx} - 1 - i\tau(x))M(dx), \end{cases}$$

where  $a, \sigma^2$  are real constants;  $M$  is a Lévy measure on  $\mathbb{R}$  characterized by the property that  $M(0) = 0$ ,  $M$  is finite outside of every neighborhood of the origin and

$$\int_1^1 \frac{x^2}{1+x^2} M(dx) < \infty;$$

the function  $\tau(x)$  is defined by

$$(12) \quad \tau(x) = \begin{cases} x & x \in I; \\ 1 & x > 1; \\ -1 & x < -1, \end{cases}$$

$I$  being the closed unit interval  $[-1,1]$ .

**3. Theorem** (cf. Thu [6], Theorem 3.2) A p.m.  $\mu \in L_\alpha, \alpha > 0$  if and only if its characteristic function  $\hat{\mu}(t)$  is of the unique form

$$(13) \quad \begin{cases} -\log \hat{\mu}(t) = iat + \sigma^2 t^2 - \\ \int_{-\infty}^{\infty} v_\alpha(x) \left( \int_0^\infty k(e^{-u}x, t) u^{\alpha-1} du \right) m(dx), \end{cases}$$

where  $m$  is a finite measure on  $\mathbb{R}$  vanishing at the origin;  $a, \sigma^2$  are real constants; the weight function  $v_\alpha(x)$  and the kernel  $k(y, t)$  are defined by

$$(14) \quad v_\alpha^{-1}(x) = \int_0^\infty \frac{e^{-2t} x^2}{1 + e^{-2t} x^2} t^{\alpha-1} dt$$

and

$$k(y, t) = \exp(i\tau(y) - 1 - i\tau(y)),$$

where the function  $\tau(y)$  is given by the formula (12). Consequently, the Lévy measure  $M$  of  $\mu$  is of the form

$$(15) \quad M(A) = \int_{-\infty}^{\infty} v_\alpha(x) \left( \int_0^\infty \chi_A(e^{-u}x) u^{\alpha-1} du \right) m(dx),$$

where  $A$  is a Borel subset of the real line separated from 0.

**4. Theorem** (cf. Thu [5]) A p.m.  $\mu$  is mixed-stable (i.e.  $\mu \in L_\infty$ ) if and only if its characteristic function  $\hat{\mu}(t)$  is of the unique form

$$(16) \quad \begin{cases} -\log \hat{\mu}(t) = iat + \sigma^2 t^2 \\ - \int_{-1}^1 \int_0^\infty q(x, t) h(x) \nu(du), \end{cases}$$

where the constants  $a, \sigma^2$  are the same as in the above theorems and  $\nu$  is a finite measure on the open interval  $(-1, 1)$ ,

$$(17) \quad \begin{cases} q(x, t) = \\ -|tx|^2|x| - itan\pi|x|signxt^{2|x|+1} & 2|x| \neq 1 \\ -|x|t - \frac{2i}{\pi}tx\log|tx| & 2|x| = 1. \end{cases}$$

Consequently, the Lévy measure of  $\mu$  is given by

$$(18) \quad M(A) = \int_{-1}^1 \int_0^\infty \chi_A(tx) \frac{dt}{t^2|x|+1} h(x) \nu(dx)$$

**2. Definition** Let  $T$  be a parameter set  $Z$  of all integers or  $R$  of all real numbers. A stochastic process  $X_t, t \in T$  is said to be *i.d.*, *stable*, *mixed-stable*,  $\alpha - s.d.$ , if for any  $t_1, t_2, \dots, t_n \in T$  and  $\lambda_1, \lambda_2, \dots, \lambda_n, n = 1, 2, \dots$  the r.v.  $\sum_1^n \lambda_j X_{t_j}$  is *i.d.*, *stable*, *mixed-stable*,  $\alpha - s.d.$ , respectively.

**3. Definition** Let  $\Lambda = \{\Lambda(A) : A \in \mathcal{S}\}$  be a real stochastic process defined on a probability space  $(\Omega, \mathcal{F}, \mathcal{P})$ , where  $\mathcal{S}$  stands for a  $\sigma$ -ring of subsets of an arbitrary non-empty set  $S$  satisfying the following condition :

There exists an increasing sequence  $S_n, n = 1, 2, \dots$  of sets in  $\mathcal{S}$  with  $\bigcap_n S_n = S$ .



We call  $\Lambda$  to be an independently scattered random measure ( or r.m. for short), if, for every sequence  $\{A_n\}$  of disjoint sets in  $\mathcal{S}$ , the random variables  $\Lambda(A_n), n = 1, 2, \dots$  are independent, and , if  $\cup_n A_n$  belongs to  $\mathcal{S}$ , then we also have

$$\Lambda(\cup_n A_n) = \sum_n \Lambda(A_n) \quad a.s.,$$

where the series is assumed to be convergent a.s. In addition, if for every  $A \in \mathcal{S}$  the distribution of  $\Lambda(A)$  is i.d., stable, mixed-stable, m.s.d. respectively, then we say that  $\Lambda$  is an i.d., stable, mixed stable,  $\alpha - s.d.r.m.$

By virtue of Theorem 2 each random variable  $\Lambda(A), A \in \mathcal{S}$  has the characteristic function

$$(19) \quad \begin{cases} -\log E \exp(it\Lambda(A)) = it\nu_0(A) + \frac{1}{2}t^2\nu_1(A) \\ - \int_{-\infty}^{\infty} (e^{itx} - 1 - it\tau(x))F_A(dx), \end{cases}$$

where  $t \in \mathbb{R}, A \in \mathcal{S}$  and  $-\infty < \nu_0(A) < \infty, 0 \leq \nu_1(A) < \infty$  and  $F_A$  is a Lévy measure on  $\mathbb{R}$ . Moreover,  $\nu_0$  is a signed measure ,  $\nu_1$  a measure and  $F_A$  a Lévy measure.

The above representation implies the following

**5.Theorem** (cf. Raiput, Rosinski [3], Proposition 2.1)  
 The characteristic function (19) can be rewritten in the unique form:

$$(20) \quad E \exp(it\Lambda(A)) = \exp\left(\int_A K(t, s)\lambda(ds)\right),$$

, where  $t \in \mathbb{R}$ ,  $A \in \mathcal{S}$  and

$$(21) \quad \begin{cases} K(t, s) = ita(s) - 1/2t^2\sigma^2(s) \\ + \int_A (e^{itx} - 1 - it\tau(x))\rho(s, dx), \end{cases}$$

with

$$(22) \quad a(s) = \frac{dv_0}{d\lambda}(s),$$

and

$$(23) \quad \sigma^2(s) = \frac{dv_1}{d\lambda}(s)$$

and  $\rho$  is given by Lemma 2.3 in [3]. Moreover, we have

$$|a(s)| + \int_{\mathbb{R}} \min\{1, x^2\}\rho(s, dx) = 1 \quad a.e.[\lambda].$$

**4.Definition** (cf. Urbanik and Woyczynski [8])

(a) If  $f$  is a simple function on  $S$ ,  $f = \sum_j x_j \chi_{A_j}$ ,  $A_j \in \mathcal{S}$  then we put, for each  $A \in \sigma(S)$

$$\int_A f d\Lambda = \sum_j \lambda(A \cap A_j);$$

(b) A measurable function  $f : (S, \sigma(S)) \rightarrow (\mathbb{R}, \mathcal{B}(\mathbb{R}))$  is said to be  $\Lambda$ -integrable if there exists a sequence  $\{f_n\}$  of simple functions as defined in (a) such that

- (i)  $f_n \rightarrow f$  a.e. $[\lambda]$ ,
- (ii) For every  $A \in \sigma(S)$ , the sequence  $\{\int_A f_n d\Lambda\}$  converges in prob., as  $n \rightarrow \infty$ . if  $f$  is  $\Lambda$ -integrable, then we put

$$\int_A f d\Lambda = P - \lim_{n \rightarrow \infty} \int_A f_n d\Lambda,$$

where  $\{f_n\}$  satisfies (i) and (ii).

Now, combining Theorems 3,4,5 we get the following:

**6.Theorem** Given  $\alpha > 0$  let  $\Lambda(A)$ ,  $A \in \mathcal{S}$  be a  $\alpha$ -s.d.r.m. Then, the characteristic function of  $\Lambda(A)$  is of the unique form (20) where

$$(24) \quad \begin{cases} K(t, s) = ita(s) - 1/2t^2\sigma^2(s) \\ + \int_A (e^{itx} - 1 - it\tau(x))\rho(s, dx), \end{cases}$$

with

$$(25) \quad a(s) = \frac{dv_0}{d\lambda}(s),$$

and

$$(26) \quad \sigma^2(s) = \frac{dv_1}{d\lambda}(s)$$

and  $\rho$  is given by Lemma 2.3 in [3]. Moreover, we have

$$|a(s)| + \int_{\mathbb{R}} \min\{1, x^2\} \rho(s, dx) = 1 \quad a.e. [\lambda].$$

**Proof** By virtue of (13) it follows that for any  $A \in \mathcal{S}$  and  $t \in \mathbb{R}$   $\Lambda(A)$  has the representation

$$(27) \quad \begin{cases} -\log E \exp(it\Lambda(A)) = it\nu_0(A) + \frac{1}{2}t^2\nu_1(A) - \\ \int_{-\infty}^{\infty} v_{\alpha}(x) \left( \int_0^{\infty} k(e^{-u}x, t) u^{\alpha-1} du \right) m(A, dx), \end{cases}$$

which, by a similar argument of Proposition 2.1 in [3], implies that there exists a unique finite measure  $\nu$  on  $\sigma(\mathcal{S}) \times \mathcal{B}(\mathcal{R})$  such that

$$\nu(A \times B) = m(A, B), \quad \text{for any } A \in \mathcal{S}, B \in \mathcal{B}(\mathcal{R}).$$

Moreover, for every  $A \in \sigma(\mathcal{S})$  we have  $\nu(A, \{0\}) = 0$ .

Now, we are in the position to present the following theorem whose proof is a simple combination of Theorem 6 and the Komogorov extension theorem and Theorem 5.2 in [3].

**7.Theorem** Given  $0 < \alpha \leq \infty$  let  $\{X_t : t \in T\}$  be an  $\alpha - s.d.$  stochastic process defined on a probability space  $(\Omega', \mathcal{P}')$ . Then there exists an  $\alpha - s.d.r.m.$ , say  $\Lambda$ , defined on the probability space  $(\Omega, \mathcal{P})$  such that

$$(\Omega = \Omega' \times [0, 1], \mathcal{P} = \mathcal{P}' \times \text{Leb}),$$

*Leb* being the Lebesgue measure on  $[0,1]$  and

$$\{X_t : t \in T\} = \left\{ \int_{\mathbb{S}} f_t(s) d\Lambda(s) : t \in T \right\},$$

where  $f_t(s) : t \in T, s \in S$  denote some measurable functions on  $\mathbb{S}$ .

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