Università degli Studi di Padova

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CBI and Hawkes processes: theory and application to power markets

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## Introduction

Energy markets and, in particular, electricity markets, exhibit very peculiar features. Electricity can be considered a commodity, just as oil and grain are; however, electricity markets are particularly different from other commodity markets.
Firstly, electricity cannot be stored economically (the problem is essentially due to the physical difficulty in storing large amount of energy and also to the fact that this procedure is extremely expensive); this implies that it is purchased primarily for consumption and then a relatively small change in demand or capacity can cause an immediate jump in price. Although producers can adjust their supply, there is always a time delay and therefore it is quite common to observe spikes (a comparatively large upward or downward movement of a price in a short period of time) in electricity prices.
Secondly, electricity consumption exhibits strong seasonality and sometimes also the capacity is affected by seasonality like hydro power generation which varies with seasons. Due to the non-storability and seasonality electricity prices are highly volatile and exhibit jumps and spikes from time to time.
Due to this properties and also to the fact that empirical evidence suggests that in many asset prices often jumps appear in cluster, to model the dynamics describing the spot price (or the future price) we need to introduce jump processes exhibiting a clustering or self-exciting behavior. There is a wide class of works in the literature with several different approaches to describe power prices evolution and a comprehensive literature review until 2008 is presented in [2].
In this work we aim to investigate if self exciting features arise in power forward prices evolution and we will focus our attention on two classes of stochastic processes: continuous branching processes with immigration (CBI) and Hawkes processes. CBI processes are commonly used in modeling population dynamics and their self exciting features describe the growth of the population due to the reproduction of the previous generations. Hawkes processes, instead, are particular extension of Poisson processes with self exciting properties, where points shows clustering effects; they have been introduced by Hawkes (1979) and the fact that they are extremely versatile makes them interesting both from a theoretical and from a practical point of view.
This work is organized in the following way: in the first chapter we set up the theoretical framework and we define and give the main properties of CBI and Hawkes processes, in the second chapter we recall some stochastic analysis notions (Girsanov Theorem) and we set up the modeling framework for forward prices based on Hawkes and CBI processes and, finally, in the third chapter we describe the algorithm to determine forward curves and then we give the parameter estimation for the model proposed. We also implement the algorithm for jump detection in forward curves and we apply it to detect jumps in historical series of data of future prices of other commodities (gold and crude oil).

## Chapter 1

## Properties of CBI and Hawkes processes

The first section of this chapter is devoted to some important definitions about random measures that will be used in the construction of the model given in Chapter 2. The references for this part are [5], [6] and [16]. The second section is devoted to the construction and the presentation of some important properties of Continuous state Branching processes (CB processes) and of Continuous state Branching processes with Immigration (CBI processes); here the main references are [20], [21] and [22]. The last section is dedicated to a review of Point processes and Poisson processes that are used to define Hawkes processes. The references for this part are [7], [19], and [26].

### 1.1 Introductory definitions

We by start describing the model which is based on stochastic differential equations (SDEs) driven by Levy random fields. First of all we set some important definitions:

Definition 1.1 (Random Measure with transition kernels). Let $(\Omega, \mathcal{F}, P)$ be a probability space and $(E, \mathcal{E})$ a measurable space. A Random Measure $X$ is a function:

$$
X: \Omega \times \mathcal{E} \rightarrow[0, \infty)
$$

such that:

- $\forall \omega \in \Omega$ the map:

$$
X(\omega, \cdot): \mathcal{E} \rightarrow[0, \infty)
$$

is a measure;

- For every $B \in \mathcal{E}$ the map:

$$
X(\cdot, B): \Omega \rightarrow[0, \infty)
$$

is $\mathcal{F}$-measurable random variable.
An alternative definition is the following:

Definition 1.2 (Random Measure as random element). Let $(\Omega, \mathcal{F}, P)$ be a probability space and $(E, \mathcal{E})$ a measurable space. A Random Measure $X$ is a measure valued element from $(\Omega, \mathcal{F}, P)$ to $(M, \mathcal{M})$ where $M$ is the space of all measures on $(E, \mathcal{E})$ and $\mathcal{M}$ is its $\sigma$-algebra.

Definition 1.3. A stochastic process $\left\{X_{t}\right\}_{t \in \mathbb{R}}$ is said to be a Lévy process if:

- $X_{0}=0$ almost surely.
- for all $t_{0}<t_{1}<\ldots<t_{n}$ the random variables $X_{t_{0}}, X_{t_{1}}-X_{t_{0}}, \ldots, X_{t_{n}}-X_{t_{n-1}}$ are mutually independent.
- for any $s<t$ the distribution of $\left(X_{t}-X_{s}\right)$ is the same of $X_{t-s}$.
- for any $\varepsilon>0$ and $h \geq 0$ it holds that:

$$
\lim _{h \rightarrow 0} P\left(\left|X_{t+h}-X_{t}\right|>\varepsilon\right)=0
$$

Finally a Lévy random field is a multi-dimensional generalization of a Lévy process
In particular we have that:
Definition 1.4 (Gaussian random measure). A Gaussian random measure is a random measure such that $\forall \omega \in \Omega$ the function:

$$
X(\omega, \cdot): \mathcal{E} \rightarrow[0, \infty)
$$

is a Gaussian measure, i.e., it is either the Dirac delta at a point $a \in \mathbb{R}$ or it has density:

$$
p\left(\cdot, a, \sigma^{2}\right): x \mapsto \frac{1}{\sqrt{2 \pi \sigma^{2}}} \exp \left(-\frac{(x-a)^{2}}{2 \sigma^{2}}\right)
$$

with respect to the Lebesgue measure where the parameters a and $\sigma$ are called mean and variance of the measure.

We consider now the case of a Gaussian random measure and we introduce the concept of white noise:

Definition 1.5. A Gaussian random measure $W: \Omega \times \mathcal{B}\left(\mathbb{R}_{+}^{2}\right) \rightarrow[0, \infty)$ is said to be $a$ white Noise on $\mathbb{R}_{+}^{2}$ if:

- $\forall A \in \mathcal{B}\left(\mathbb{R}_{+}^{2}\right)$ for which $\lambda^{2}(A)<\infty\left(\lambda^{2}\right.$ denotes the Lebesgue measure on $\left.\mathbb{R}^{2}\right)$ we have that $W(\cdot, A)$ is a Gaussian random variable with zero mean and variance equal to $\lambda^{2}(A)$.
- Given $A_{1}, \ldots, A_{n}$ disjoint Borel set with finite Lebesgue measure, then the random variables

$$
W\left(\cdot, A_{1}\right), \ldots, W\left(\cdot, A_{n}\right)
$$

are mutually independent.
We also need the following:

Definition 1.6. We say that $N: \Omega \times \mathcal{B}\left(\mathbb{R}_{+}^{3}\right) \rightarrow[0, \infty)$ is a Poisson random measure with intensity $\lambda$ (with $\lambda$ Borel measure on $\mathbb{R}_{+}^{3}$ ) if it is the product measure of the Lebesgue measure on $\mathbb{R}_{+} \times \mathbb{R}_{+}$with a Borel measure $\mu$ on $\mathbb{R}_{+}$such that:

$$
\int_{0}^{\infty}\left(z \wedge z^{2}\right) \mu(d z)<\infty
$$

In particular the Borel measure $\mu$ in Definition (1.6) is a Levy measure since:

$$
1 \wedge z^{2} \leq\left(z \wedge z^{2}\right) \mathbf{1}_{(0,1]}(z)+\left(z \wedge z^{2}\right) \mathbf{1}_{(1, \infty)}(z)=z \wedge z^{2}
$$

This implies that:

$$
\int_{0}^{\infty}\left(1 \wedge z^{2}\right) \mu(d z) \leq \int_{0}^{\infty}\left(z \wedge z^{2}\right) \mu(d z)<\infty
$$

We also introduce:
Definition 1.7. Let $N$ be a Poisson random measure, we define the compensated Poisson random measure of intensity $\lambda$ as $\widetilde{N}=N-\lambda$.

At this point we introduce the filtration $\mathbb{F}=\left\{\mathcal{F}_{t}\right\}_{t \geq 0}$ as the natural filtration generated by Lévy's random field, namely for any Borel subset $A \in \mathcal{B}\left(\mathbb{R}_{+}\right)$and $B \in \mathcal{B}\left(\mathbb{R}^{2}\right)$ of finite Lebesgue measure the processes: $(W([0, t] \times A), t \geq 0)$ and $(\tilde{N}([0, t] \times B), t \geq 0)$ are $P$-Martingales.
At this point we are ready to set up the stochastic differential equation that we are going to use: we set $a, b, \sigma, \gamma \in \mathbb{R}_{+}$positive constants and:

- $W(d s, d u)$ a white noise on $\mathbb{R}^{2}$ with unit covariance matrix.
- $\widetilde{N}(d s, d u, d z)$ an independent compensated Poisson random measure on $\mathbb{R}^{3}$ with intensity $d s d u \mu(d z)$, with $\mu$ Lévy measure on $\mathbb{R}^{+}$satisfying $\int_{0}^{\infty}\left(z \wedge z^{2}\right) \mu(d z)<\infty$.


### 1.2 Laplace transform of a measure

We start recalling some basic properties of the Laplace transform of a measure and then we continue defining the continuous state branching process (CB) and continuous state branching process with immigration (CBI). Let $\mathcal{B}([0, \infty))$ be the Borel $\sigma$-algebra on the positive half line and let $B_{b}([0, \infty))$ the set of bounded measurable functions on $[0, \infty)$. Given a finite measure $\mu$ on $[0, \infty)$ we define the Laplace transform $L_{\mu}$ of $\mu$ by:

$$
\begin{equation*}
L_{\mu}(\lambda)=\int_{[0, \infty)} e^{-\lambda x} \mu(d x) \quad \lambda \geq 0 \tag{1.1}
\end{equation*}
$$

Theorem 1.1. A finite measure on $[0, \infty)$ is uniquely determined by its Laplace transform.
Proof. Suppose that $\mu_{1}$ and $\mu_{2}$ are finite measures on $[0, \infty)$ and that $L_{\mu_{1}}(\lambda)=L_{\mu_{2}}(\lambda)$ for all $\lambda \geq 0$. Then set $\mathcal{H}=\left\{x \mapsto e^{-\lambda x}: \lambda \geq 0\right\}$ and $\mathcal{L}$ the set of functions in $B_{b}([0, \infty))$ such that:

$$
\int_{[0, \infty)} F(x) \mu_{1}(d x)=\int_{[0, \infty)} F(x) \mu_{2}(d x)
$$

Then $\mathcal{H}$ is closed under multiplication and $\mathcal{L}$ is a monotone vector space containing $\mathcal{H}$. Moreover we have that $\sigma(\mathcal{H})=\mathcal{B}([0, \infty))$ and hence we have that by the monotone class theorem:

$$
\mathcal{L} \supset \sigma_{b}(\mathcal{H})=B_{b}([0, \infty))
$$

which proves the result.
We state now a theorem about converging result of Laplace transform of a measure.
Theorem 1.2. Let $\left\{\mu_{n}\right\}_{n \in \mathbb{N}}$ be a sequence of finite measures on $[0, \infty)$ and $\lambda \mapsto L(\lambda)$ be a continuous function on $[0, \infty)$. If $\lim _{n \rightarrow \infty} L_{\mu_{n}}(\lambda)=L(\lambda)$ for every $\lambda \geq 0$ then there is a finite measure $\mu$ such that $L_{\mu}=L$ and $\lim _{n \rightarrow \infty} \mu_{n}=\mu$ by weak convergence.

Corollary 1.1. Let $\mu_{1}, \ldots, \mu_{n}$ be finite measures on $[0, \infty)$. Then $\mu_{n} \rightarrow \mu$ weakly if and only if $L_{\mu_{n}}(\lambda) \rightarrow L(\lambda)$.

Proof. If $\mu_{n} \rightarrow \mu$ weakly we have that $\lim _{n \rightarrow \infty} L_{\mu_{n}}(\lambda)=L_{\mu}(\lambda)$ for every $\lambda \geq 0$, indeed being $x \mapsto e^{-\lambda x}$ continuous and bounded for every $\lambda \geq 0$ we have that:

$$
\lim _{n \rightarrow \infty} L_{\mu_{n}}(\lambda)=\lim _{n \rightarrow \infty} \int_{0}^{\infty} e^{-\lambda x} \mu_{n}(d x)=\int_{0}^{\infty} e^{-\lambda x} \mu(d x)=L_{\mu}(\lambda)
$$

The converse is a consequence of the Theorem 1.2.
We conclude this section adding some additional results:

- given two probability measures $\mu_{1}, \mu_{2}$ we define $\mu_{1} \times \mu_{2}$ their product measure on $[0, \infty)^{2}$.
- The image of $\mu_{1} \times \mu_{2}$ under the mapping $\left(x_{1}, x_{2}\right) \mapsto x_{1}+x_{2}$ is called convolution of $\mu_{1}$ and $\mu_{2}$ and it is denoted by $\mu_{1} * \mu_{2}$ and it is a probability measure on $[0, \infty)$

For any $F \in_{b} B([0, \infty))$ we have that:

$$
\begin{equation*}
\int_{[0, \infty)} F(x)\left(\mu_{1} * \mu_{2}\right)(d x)=\int_{[0, \infty)} \mu_{1}(d x) \int_{[0, \infty)} F\left(x_{1}+x_{2}\right) \mu_{2}(d x) \tag{1.2}
\end{equation*}
$$

If $X_{1}$ and $X_{2}$ are random variables with distributions $\mu_{1}$ and $\mu_{2}$ then $X_{1}+X_{2}$ is a random variable with distribution $\mu_{1} * \mu_{2}$. Finally we have that:

$$
\begin{equation*}
L_{\mu_{1} * \mu_{2}}(\lambda)=L_{\mu_{1}}(\lambda) L_{\mu_{2}}(\lambda) \quad \lambda \geq 0 \tag{1.3}
\end{equation*}
$$

Let now:

$$
\begin{aligned}
& \mu^{* 0}=\delta_{0} \\
& \mu^{* n}=\mu^{*(n-1)} * \mu \quad \text { for } n \geq 1
\end{aligned}
$$

We say that a probability distribution $\mu$ on $[0, \infty)$ is infinitely divisible if for each integer $n \geq 1$ there is a probability $\mu_{n}$ such that $\mu=\mu_{n}^{* n}$; in this case $\mu_{n}$ is called the n-th root of $\mu$.
A positive random variable $X$ is said to be infinitely divisible if it admits infinitely divisible distribution on $[0, \infty)$.

### 1.3 Construction of a CB process

CB and CBI processes are a particular class of stochastic processes commonly used in modelling population dynamics. The self-exciting features describe the growth of the population due to the reproduction of the previous generations. Our aim is to use them to describe jumps generated by previous jumps. In this section we present a possible construction of a CB-Process.

### 1.3.1 Galton Watson Branching process

Let $\{p(j): j \in \mathbb{N}\}$ be a probability distribution of a r.v. $X$ on the space of positive integers $\mathbb{N}$. It is well known that $\{p(j): j \in \mathbb{N}\}$ is uniquely determined by its generating function $g: \mathbb{N} \rightarrow \mathbb{R}$ defined by:

$$
g(z)=E\left(z^{X}\right)=\sum_{j=0}^{\infty} p(j) z^{j} \quad|z| \leq 1
$$

At this point we suppose to have $\left\{\xi_{n, i}: n, i=1,2, \ldots\right\}$ a family of $\mathbb{N}$-valued random variables with distribution $\{p(j): j \in \mathbb{N}\}$. Given an $\mathbb{N}$-valued random variable $Z_{0}$ independent on $\left\{\xi_{i, n}\right\}$ we define inductively for $n=1,2, \ldots$ :

$$
\begin{equation*}
Z_{n}=\sum_{i=1}^{Z_{n-1}} \xi_{n, i} \quad Z_{0}=0 \tag{1.4}
\end{equation*}
$$

Of course we set $\sum_{i=1}^{0}=0$. For each $i \in \mathbb{N}$ we denote with $\{Q(i, j): j \in \mathbb{N}\}$ the $i$-fold convolution of $\{p(j): j \in \mathbb{N}\}$, that is, $Q(i, j)=p^{* i}(j)$ for $i, j \in \mathbb{N}$. For any $n \geq 1$ and $\left\{i_{0}, i_{1}, i_{2}, \ldots, i_{n-1}=i, j\right\} \subset \mathbb{N}$ we have that:

$$
\begin{aligned}
P\left(Z_{n}=j \mid Z_{0}=i_{0}, \ldots, Z_{n-1}=i\right) & =P\left(\sum_{i=1}^{Z_{n-1}} \xi_{n, i}=j \mid \sum_{i=0}^{Z_{n-2}} \xi_{n-1, i}=i_{n-1}, \ldots, Z_{0}=i_{0}\right) \\
& =P\left(\sum_{i=1}^{Z_{n-1}} \xi_{n, i}=j \mid \sum_{i=0}^{Z_{n-2}} \xi_{n-1, i}=i_{n-1}\right) \\
& =P\left(\sum_{k=1}^{i} \xi_{n, k}=j\right)=p^{* i}(j)=Q(i, j) .
\end{aligned}
$$

Then $Z_{n}: n \in \mathbb{N}$ is an $\mathbb{N}$-valued Markov chain with one step transition matrix $Q=$ $Q(i, j): i, j \in \mathbb{N}$; in particular the random variable $Z_{n}$ can be tought as the number of individuals in the generation $n$ of an evolving population system. After one time unit each individual of the population splits independently on the others into a random number of offspring according to the distribution $\{p(j): j \in \mathbb{N}\}$. Clearly we have for $i \in \mathbb{N}$ and for $|z| \leq 1$ :

$$
\begin{equation*}
\sum_{j=0}^{\infty} Q(i, j) z^{j}=\sum_{j=0}^{\infty} p^{* i}(j) z^{j}=g(z)^{i} \tag{1.5}
\end{equation*}
$$

Moreover the transition matrix $Q$ satisfies the branching property:

$$
\begin{equation*}
Q\left(i_{1}+i_{2}, \cdot\right)=Q\left(i_{1}, \cdot\right) * Q\left(i_{2}, \cdot\right) \tag{1.6}
\end{equation*}
$$

Namely, that different individuals of the population propagate independently on each other. A Markov chain with state space $\mathbb{N}$ with one step transition matrix defined by (1.5) is called a Galton Watson branching process (GW-process) or a Bienaymé-GaltonWatson branching process with branching distribution given by $g$.
By a general result on the theory of Markov chains we have that for $n \geq 1$ the $n$-step transition matrix of the $G W$-process is just the $n$-fold product matrix $Q^{n}=\left\{Q^{n}(i, j)\right.$ : $i, j \in \mathbb{N}\}$.

Proposition 1.1. For $n \geq 1$ and $i \in \mathbb{N}$ we have:

$$
\begin{equation*}
\sum_{j=0}^{\infty} Q^{n}(i, j) z^{j}=g^{\circ n}(z)^{i} \tag{1.7}
\end{equation*}
$$

where $g^{\circ n}(z)$ is defined iteratively by: $g^{\circ 0}(z)=z$ and:

$$
g^{\circ n}=g \circ g^{\circ(n-1)}(z)=g\left(g^{\circ(n-1)}(z)\right) .
$$

Proof. We know that the previous property holds for $n=1$. Suppose now that it holds for some $n \geq 1$, then we have that:

$$
\begin{aligned}
\sum_{j=0}^{\infty} Q^{n+1}(i, j) z^{j} & =\sum_{j=0}^{\infty} \sum_{k=0}^{\infty} Q(i, k) Q^{n}(k, j) z^{j} \\
& =\sum_{k=0}^{\infty} Q(i, k) g^{\circ n}(z)^{i}
\end{aligned}
$$

And this proves inductively the thesis.
It is easy to see that 0 is a trap for the $G W$-process; indeed if at some $n \geq 1$ we have that $Z_{n}=0$, then:

$$
Z_{n+1}=\sum_{i=1}^{Z_{n}} \xi_{n, i}=0
$$

And therefore by induction on $n$ we conclude that $Z_{k}=0 \forall k \geq n$.
If $g^{\prime}(-1)<\infty$ by differentiating both sides of (1.7) we see that the first moment of the distribution $\left\{Q^{n}(i, j): j \in \mathbb{N}\right\}$ is given by:

$$
\begin{equation*}
\sum_{j=1}^{\infty} j Q^{n}(i, j)=i g^{\prime}(1-)^{n} \tag{1.8}
\end{equation*}
$$

### 1.3.2 Continuous extension of a Galton-Watson process

Suppose now to have a sequence of GW-processes $\left\{X_{k}(n): n \geq 0\right\}_{k \geq 1}$ with branching distribution given by the probability generating function $g_{k}$, for $k=1,2, \ldots$. Consider now:

$$
\begin{equation*}
Z_{k}(n)=\frac{1}{k} X_{k}(n) \quad n \geq 0 \tag{1.9}
\end{equation*}
$$

the process defined by (1.9) is a Markov chain with state space $E_{k}=\left\{0, k^{-1}, 2 k^{-1}, \ldots\right\}$ and $n$-step transition probability given by $Q_{k}^{n}(x, d y)$ determined by:

$$
\begin{equation*}
\int_{E_{k}} e^{\lambda y} Q_{k}^{n}(x, d y)=g_{k}^{\circ n}\left(e^{-\lambda / k}\right)^{k x} \tag{1.10}
\end{equation*}
$$

Suppose now that $\left\{\gamma_{k}\right\}_{k}$ is a positive sequence $\gamma_{k} \rightarrow \infty$ as $k \rightarrow \infty$ and let $\left\lfloor\gamma_{k} t\right\rfloor$ denote the integer part of $\gamma_{k} t$. Given $Z_{k}(0)=x$ we know that the random variable $Z_{k}\left(\left\lfloor\gamma_{k} t\right\rfloor\right)=$ $k^{-1} X_{k}\left(\left\lfloor\gamma_{k} t\right\rfloor\right)$ has distribution $Q_{k}^{\left\lfloor\gamma_{k} t\right\rfloor}(x, \cdot)$ on $E_{k}$ determined by:

$$
\begin{equation*}
\int_{E_{k}} e^{-\lambda y} Q_{k}^{\left\lfloor\gamma_{k} t\right\rfloor}(x, d y)=\exp \left\{-x v_{k}(t, \lambda)\right\} \tag{1.11}
\end{equation*}
$$

where:

$$
\begin{equation*}
v_{k}(t, \lambda)=-k \log \left(g_{k}^{\circ\left\langle\gamma_{k} t\right\rfloor}\left(e^{-\lambda / k}\right)\right) \tag{1.12}
\end{equation*}
$$

We are interested in the asymptotic behavior of the sequence of continuous time processes: $\left\{Z_{k}\left(\left\lfloor\gamma_{k} t\right\rfloor\right): t \geq 0\right\}$ as $k \rightarrow \infty$. By (1.12) for $\gamma_{k}^{-1}(i-1) \leq t \leq \gamma_{k}^{-1} i$ we have:

$$
v_{k}(t, \lambda)=v_{k}\left(\gamma_{k}^{-1}\left\lfloor\gamma_{k} t\right\rfloor, \lambda\right)=v_{k}\left(\gamma_{k}^{-1}(i-1), \lambda\right)
$$

It follows that (the following chain of calculations is taken from [20]):

$$
\begin{aligned}
v_{k}(t, \lambda) & =v_{k}(0, \lambda)+\sum_{j=1}^{\left\lfloor\gamma_{k} t\right\rfloor}\left[v_{k}\left(\gamma_{k}^{-1} j, \lambda\right)-v_{k}\left(\gamma_{k}^{-1}(j-1), \lambda\right)\right] \\
& =\lambda-k \sum_{j=1}^{\left\lfloor\gamma_{k} t\right\rfloor}\left[\log \left(g_{k}^{\circ j}\left(e^{-\lambda / k}\right)\right)-\log \left(g_{k}^{\circ(j-1)}\left(e^{-\lambda / k}\right)\right)\right] \\
& =\lambda-k \sum_{j=1}^{\left\lfloor\gamma_{k} t\right\rfloor} \log \left[\left(g_{k}\left(g_{k}^{\circ(j-1)}\left(e^{-\lambda / k}\right)\right)\left(g_{k}^{\circ(j-1)}\left(e^{-\lambda / k}\right)\right)^{-1}\right]\right. \\
& =\lambda-\gamma_{k}^{-1} \sum_{j=1}^{\left\lfloor\gamma_{k} t\right\rfloor} \varphi_{k}\left(-k \log g_{k}^{\circ(j-1)}\left(e^{-\lambda / k}\right)\right) \\
& =\lambda-\gamma_{k}^{-1} \sum_{j=1}^{\left\lfloor\gamma_{k} t\right\rfloor} \varphi_{k}\left(v_{k}\left(\gamma_{k}^{-1}(j-1)\right)\right) \\
& =\lambda-\int_{0}^{\gamma_{k}^{-1}\left\lfloor\gamma_{k} t\right\rfloor} \varphi_{k}\left(v_{k}(s, \lambda)\right) d s
\end{aligned}
$$

where:

$$
\begin{equation*}
\varphi_{k}(z)=k \gamma_{k} \log \left[g_{k}\left(e^{-z / k}\right) e^{z / k}\right] \tag{1.13}
\end{equation*}
$$

We can rewrite the previous function as:

$$
\begin{equation*}
\varphi_{k}(z)=k \gamma_{k} \log \left[1+\left(k \gamma_{k}\right)^{-1} \tilde{\varphi}_{k}(z) e^{z / k}\right] \tag{1.14}
\end{equation*}
$$

where:

$$
\begin{equation*}
\tilde{\varphi}_{k}(z)=k \gamma_{k}\left[g_{k}\left(e^{-z / k}\right)-e^{-z / k}\right] \tag{1.15}
\end{equation*}
$$

We state now a lemma which shows that the two sequences (1.13) and (1.15) are not very different:

Lemma 1.1. Suppose that the sequence $\left\{\varphi_{k}\right\}_{k \in \mathbb{N}}$ is given by (1.13) and that $\left\{\tilde{\varphi}_{k}\right\}_{k \in \mathbb{N}}$ is given by (1.15). Assume moreover that $\left\{\tilde{\varphi}_{k}\right\}_{k \in \mathbb{N}}$ is bounded on each bounded interval. an Then we have:

- $\lim _{k \rightarrow \infty}\left|\tilde{\varphi}_{k}(z)-\varphi_{k}(z)\right|=0$ uniformly on each bounded interval.
- $\left\{\varphi_{k}\right\}_{k}$ is uniformly Lipschitz on each bounded interval if and only if so is $\left\{\tilde{\varphi}_{k}\right\}_{k}$

This lemma allow us to deduce that if either $\left\{\varphi_{k}\right\}_{k}$ or $\left\{\tilde{\varphi}_{k}\right\}_{k}$ is uniformly Lipschitz on each bounded interval then they converge or diverge simultaneously and in the convergent case they have the same limit. For convenience we formulate the following condition:

Assumption 1.1. The sequence $\left\{\tilde{\varphi}_{k}\right\}_{k}$ is uniformly Lipschitz on $[0, a]$ for every $a \geq 0$ and there is a function $\varphi$ on $[0, \infty)$ so that $\tilde{\varphi}_{k}(z) \rightarrow \varphi(z)$ uniformly on $[0, a]$ as $k \rightarrow \infty$.

Proposition 1.2. Suppose that Assumption 1.1 is satisfied, then the limit function of the $\lim _{k \rightarrow \infty} \varphi_{k}=\varphi$ has representation:

$$
\begin{equation*}
\varphi(z)=b z+c z^{2}+\int_{0}^{\infty}\left(e^{-z u}-1+z u\right) m(d u), \quad z \geq 0 \tag{1.16}
\end{equation*}
$$

where $c \geq 0, b$ is a real constant and $m(d u)$ is a $\sigma$-finite measure such that:

$$
\int_{[0, \infty)}\left(u \wedge u^{2}\right) m(d u)<\infty
$$

Proof. For each $k \geq 1$ we define the function $\varphi_{k}$ on $[0, k]$ by:

$$
\begin{equation*}
\varphi_{k}(z)=k \gamma_{k}\left[g_{k}(1-z / k)-(1-z / k)\right] . \tag{1.17}
\end{equation*}
$$

From (1.15) and (1.17) we have:

$$
\tilde{\varphi}_{k}^{\prime}(z)=\gamma_{k} e^{-z / k}\left[1-g_{k}^{\prime}\left(e^{-z / k}\right)\right] \quad z \geq 0
$$

and:

$$
\varphi_{k}^{\prime}(z)=\gamma_{k}\left[1-g_{k}^{\prime}(1-z / k)\right] \quad 0 \leq z \leq k .
$$

Since $\left\{\tilde{\varphi}_{k}\right\}_{k}$ is uniformly Lipschitz on each bounded interval the sequence $\left\{\tilde{\varphi}_{k}\right\}_{k}$ is uniformly bounded on each bounded interval too. Then also $\left\{\varphi_{k}^{\prime}\right\}_{k}$ is uniformly bounded on each bounded interval and so the sequence $\left\{\varphi_{k}\right\}_{k}$ is uniformly Lipschitz on each bounded interval. Let now $a \geq 0$, for $k \geq a$ and $0 \leq z \leq a$, by the mean value theorem we have:

$$
\begin{aligned}
\tilde{\varphi}_{k}(z)-\varphi_{k}(z) & =k \gamma_{k}\left[g_{k}\left(e^{-z / k}\right)-g_{k}\left(1-\frac{z}{k}\right)-e^{-z / k}+\left(1-\frac{z}{k}\right)\right] \\
& =k \gamma_{k}\left[g_{k}^{\prime}\left(\eta_{k}\right)-1\right]\left(e^{-z / k}-1+\frac{z}{k}\right)
\end{aligned}
$$

where:

$$
1-\frac{a}{k} \leq 1-\frac{z}{k} \leq \eta_{k} \leq e^{-z / k} \leq 1
$$

Choose now $k_{0} \geq a$ so that $e^{-2 a / k_{0}} \leq 1-\frac{a}{k_{0}}$. Then $e^{-2 a / k} \leq 1-\frac{a}{k}$ for all $k \geq k_{0}$ and hence:

$$
\gamma_{k}\left|g_{k}^{\prime}\left(\eta_{k}\right)-1\right| \leq \sup _{0 \leq z \leq 2 a} \gamma_{k}\left|g_{k}^{\prime}\left(e^{-z / k}\right)-1\right|=\sup _{0 \leq z \leq 2 a} e^{-z / k}\left|\tilde{\varphi}_{k}^{\prime}(z)\right|
$$

Since $\left\{\tilde{\varphi}_{k}^{\prime}\right\}$ is uniformly bounded on $[0,2 a]$ the sequence $\left\{\gamma_{k}\left|g_{k}^{\prime}\left(\eta_{k}\right)-1\right|: k \geq k_{0}\right\}$ is bounded. Then $\lim _{k \rightarrow \infty}\left|\tilde{\varphi}_{k}(z)-\varphi_{k}(z)\right|=0$ uniformly on each bounded interval and this implies that:

$$
\lim _{k \rightarrow \infty} \varphi_{k}(z)=\varphi(z)
$$

uniformly on each bounded interval. The result follows from Corollary 1.46 in [21]
Proposition 1.3. For any function $\varphi$ with representation (1.16) there is a sequence $\left\{\tilde{\varphi}_{k \in \mathbb{N}}\right\}_{k}$ satisfying Assumption 1.1.
Proof. By the proof of proposition 1.2 it suffices to construct a sequence $\left\{\varphi_{k}\right\}_{k}$ via expression (1.17) that is uniformly Lipschitz on $[0, a]$ and $\varphi_{k}(z) \rightarrow \varphi(z)$ uniformly on $[0, a]$ for every $a \geq 0$. To simplify the formulations we decompose the function $\varphi$ into two parts. Let $\varphi_{0}(z)=\varphi(z)-b z$. We first define:

$$
\gamma_{0, k}=(1+2 c) k+\int_{(0, \infty)} u\left(1-e^{-k u}\right) m(d u)
$$

and:

$$
g_{0, k}(z)=z+k^{-1} \gamma_{0, k}^{-1} \varphi_{0}(k(1-z)) \quad|z| \leq 1 .
$$

The function $z \mapsto g_{0, k}(z)$ is analytic, satisfies $g_{0, k}(1)=1$ and:

$$
\frac{d^{n}}{d z^{n}} g_{0, k}(0) \geq 0 \quad n \geq 0
$$

Therefore $g_{0, k}(\cdot)$ is a probability generating function. Let $\varphi_{0, k}$ be defined by (1.17) with the pair $\left(\gamma_{k}, g_{k}\right)$ replaced by $\left(\gamma_{0, k}, g_{0, k}\right)$. Then $\varphi_{0, k}(z)=\varphi_{0}(z)$ if $0 \leq z \leq k$ and this completes the proof if $b=0$.
If $b \neq 0$ we set:

$$
g_{1, k}(z)=\frac{1}{2}\left(1+\frac{b}{|b|}\right)+\frac{1}{2}\left(1-\frac{b}{|b|}\right) z^{2} .
$$

Let $\gamma_{1, k}=|b|$ and let $\varphi_{1, k}(z)$ defined by (1.17) with the pair ( $\gamma_{k}, g_{k}$ ) replaced by $\left(\gamma_{1, k}, g_{1, k}\right)$. Then:

$$
\varphi_{1, k}(z)=b z+\frac{1}{2 k}(|b|-b) z^{2}
$$

Finally let $\gamma_{k}=\gamma_{0, k}+\gamma_{1, k}$ and:

$$
g_{k}=\gamma_{k}^{-1}\left(\gamma_{0, k} g_{0, k}+\gamma_{1, k} g_{1, k}\right)
$$

Then the sequence $\varphi_{k}(z)$ defined by (1.17) is equal to $\varphi_{0, k}(z)+\varphi_{1, k}(z)$ which satisfies the required condition.

Lemma 1.2. Suppose that the sequence $\left\{\tilde{\varphi}_{k}\right\}_{k \in \mathbb{N}}$ defined by (1.15) is uniformly Lipschitz on $[0,1]$. Then there exist constants $B, N \geq 0$ such that

$$
v_{k}(t, \lambda) \leq \lambda e^{B t}
$$

for every $t, \lambda \geq 0$ and $k \geq N$.

Proof. Let $b_{k}:=\varphi_{k}^{\prime}(0+)$ for $k \geq 1$. Since $\left\{\tilde{\varphi}_{k}\right\}_{k}$ is uniformely Lipschitz on $[0,1]$ the sequence $\left\{b_{k}\right\}_{k}$ is bounded. From (1.15) we have that $b_{k}=\gamma_{k}\left[1-g_{k}^{\prime}(1-)\right]$. Then:

$$
\int_{E_{k}} y Q_{k}^{\left\lfloor\gamma_{k} t\right\rfloor}(x, d y)=x g_{k}^{\prime}(1-)^{\left\lfloor\gamma_{k} t\right\rfloor}=x\left(1-\frac{b_{k}}{\gamma_{k}}\right)^{\left\lfloor\gamma_{k} t\llcorner \right.}
$$

Let $B$ a constant such that $2\left|b_{k}\right| \leq B$ for all $k \geq 1$. Since $\gamma_{k} \rightarrow \infty$ as $k \rightarrow \infty$ there is $N \geq 1$ such that:

$$
0 \leq\left(1-\frac{b_{k}}{\gamma_{k}}\right)^{\gamma_{k} / B} \leq\left(1+\frac{B}{2 \gamma_{k}}\right)^{\gamma_{k} / B} \leq e \quad k \geq N .
$$

It follows that for $t \geq 0$ and $k \geq N$ :

$$
\int_{E_{k}} y Q_{k}^{\left\lfloor\gamma_{k} t\right\rfloor}(x, d y) \leq x \exp \left\{B\left\lfloor\gamma_{k} t\right\rfloor / \gamma_{k}\right\} \leq x e^{B t}
$$

The estimate required can be obtained from (1.8) and Jensen's inequality.
Theorem 1.3. Suppose that Assumption 1.1 holds. Then for every $a \geq 0$ we have that $v_{k}(t, \lambda) \rightarrow v_{t}(\lambda)$ uniformly on $[0, a]^{2}$ and the limit function solves the integral equation:

$$
\begin{equation*}
v_{t}(\lambda)=\lambda-\int_{0}^{t} \varphi\left(v_{s}(\lambda)\right) d s \quad \lambda, t \geq 0 \tag{1.18}
\end{equation*}
$$

Proof. Recall that:

$$
v_{k}(t, \lambda)=\lambda-\int_{0}^{\gamma_{k}^{-1}\left\lfloor\gamma_{k} t\right\rfloor} \varphi_{k}\left(v_{k}(s, \lambda)\right) d s
$$

so that we can write:

$$
\begin{equation*}
v_{k}(t, \lambda)=\lambda+\varepsilon_{k}(t, \lambda)-\int_{0}^{t} \varphi_{k}\left(v_{k}(s, \lambda)\right) d s \tag{1.19}
\end{equation*}
$$

where:

$$
\varepsilon_{k}(t, \lambda)=\left(t-\gamma_{k}^{-1}\left\lfloor\gamma_{k} t\right\rfloor\right) \varphi_{k}\left(v_{k}\left(\gamma_{k}^{-1}\left\lfloor\gamma_{k} . t\right\rfloor, \lambda\right)\right)
$$

By lemma 1.1 and Assumption 1.1 for any $0 \leq \varepsilon \leq 1$ we can choose $N \geq 1$ so that $\left|\varphi_{k}(z)-\varphi(z)\right| \leq \varepsilon$ for $k \geq N$ and $0 \leq z \leq a e^{B a}$. It follows that for $0 \leq t \leq a$ and $0 \leq \lambda \leq a$ :

$$
\begin{equation*}
\left|\varepsilon_{k}(t, \lambda)\right| \leq \gamma_{k}^{-1}\left|\varphi_{k}\left(v_{k}\left(\gamma_{k}^{-1}\left\lfloor\gamma_{k} t\right\rfloor, \lambda\right)\right)\right| \leq \gamma_{k}^{-1}\left(1+\sup _{0 \leq z \leq a e^{B a}}|\varphi(z)|\right)=\gamma_{k}^{-1} M \tag{1.20}
\end{equation*}
$$

For $n \geq K \geq N$ let:

$$
K_{k, n}(t, \lambda)=\sup _{0 \leq s \leq t}\left|v_{n}(s, \lambda)-v_{k}(s, \lambda)\right| .
$$

By (1.19) and (1.20) we obtain for $(t, \lambda) \in[0, a]^{2}$ :

$$
\begin{aligned}
K_{n}(t, \lambda) & \leq 2 \gamma_{k}^{-1} M+\int_{0}^{t} \mid \varphi_{k}\left(v_{k}(s, \lambda)-\varphi_{n}\left(v_{n}(s, \lambda)\right) \mid d s\right. \\
& \leq 2\left(\gamma_{k}^{-1} M+\varepsilon a\right)+\int_{0}^{t} \mid \varphi_{k}\left(v_{k}(s, \lambda)-\varphi_{n}\left(v_{n}(s, \lambda)\right) \mid d s\right. \\
& \leq 2\left(\gamma_{k}^{-1} M+\varepsilon a\right)+L \int_{0}^{t} K_{k, n}(s, \lambda) d s
\end{aligned}
$$

where $L=\sup _{0 \leq z \leq a e^{B a}}\left|\varphi^{\prime}(z)\right|$. By Gronwall's ${ }^{1}$ inequality:

$$
K_{k, n}(t, \lambda) \leq 2\left(\gamma_{k}^{-1} M+\varepsilon a\right) \exp L t \quad 0 \leq t, \lambda \leq a
$$

Then $v_{k}(t, \lambda) \rightarrow v_{t}(\lambda)$ uniformly on $[0, a]^{2}$ as $k \rightarrow \infty$ and from (1.19) we get the thesis
The following theorem (whose proof can be founded on [20]) shows some important properties of the function $v_{t}(\lambda)$ :

Theorem 1.4. Suppose that $\varphi$ is a function given by (1.16). Then for any $\lambda \geq 0$ there is a unique positive solution $t \mapsto v_{t}(\lambda)$ to (1.18); moreover the solution satisfies the semigroup property:

$$
\begin{equation*}
v_{r+t}(\lambda)=v_{r} \circ v_{t}(\lambda) \tag{1.21}
\end{equation*}
$$

Theorem 1.5. Suppose that $\varphi$ is a function given by (1.2). For any $\lambda \geq 0$ let $t \mapsto v_{t}(\lambda)$ be the unique positive solution of (1.18). Then we define a transition semigroup $\left(Q_{t}\right)_{t \geq 0}$ on $[0, \infty)$ by:

$$
\begin{equation*}
\int_{[0, \infty)} e^{-\lambda y} Q_{t}(x, d y)=e^{-x v_{t}(\lambda)} \quad \lambda \geq 0, x \geq 0 \tag{1.22}
\end{equation*}
$$

Proof. There is a sequence $\left\{\tilde{\varphi}_{k}\right\}_{k}$ in form (1.15) satisfying Assumption 1.1. By Theorem 1.3 we have that $v_{k}(t, \lambda) \rightarrow v_{t}(\lambda)$ uniformly on $[0, a)^{2}$ as $k \rightarrow \infty$ for every $a \geq 0$. Taking $x_{k} \in E_{k}$ such that $x_{k} \rightarrow x$ as $k \rightarrow \infty$ we see by theorem 1.2 that (1.22) defines a probability measure $Q_{t}(x, d y)$ on $[0, \infty)$ and that:

$$
\lim _{k \rightarrow \infty} Q_{k}^{\left\lfloor\gamma_{k} t\right\rfloor}\left(x_{k}, \cdot\right)=Q_{t}(x, \cdot)
$$

by weak convergence. By a monotone class argument we can see that $Q_{t}(x, d y)$ is a kernel on $[0, \infty)$. The semigroup property of the family of kernels follows from (1.21) and (1.22).

Proposition 1.4. For every $t \geq 0$ the function $\lambda \mapsto v_{t}(\lambda)$ is strictly increasing on $[0, \infty)$
Proof. By the continuity of $t \mapsto v_{t}(\lambda)$, for any $\lambda_{0}>0$ there is $t_{0}>0$ so that $v_{t}\left(\lambda_{0}\right)>0$ for $0 \leq t \leq t_{0}$. Then (1.22) implies $Q_{t}(x,\{0\})<1$ for $x>0$ and $0 \leq t \leq t_{0}$, and so $\lambda \mapsto v_{t}(\lambda)$ is strictly increasing for $0 \leq t \leq t_{0}$. By the semigroup property (1.21) we infer $\lambda \mapsto v_{t}(\lambda)$ is strictly increasing for all $t \geq 0$.

Assume now that $E$ is a metrizable locally compact topological space. We also assume that $E$ is countable at infinity, meaning that $E$ is a countable union of compact sets. Let $C_{0}(E)$ the space of real valued continuous functions that vanishes at infinity (or uniformly smaller than $\varepsilon$ ) outside of a compact set); this space is a Banach space with the standard supremum norm.

Definition 1.8. Let $\left\{Q_{t}\right\}_{t \geq 0}$ a transition semigroup on $E$, we say that $\left\{Q_{t}\right\}_{t \geq 0}$ is a Feller semigroup if:

[^0]- $Q_{t} f \in C_{0}(E)$ for all $f \in C_{0}(E)$;
- $\lim _{t \rightarrow 0}\left\|Q_{t} f-f\right\|=0$ for all $f \in C_{0}(E)$.

A Markov process on $E$ is a Feller process if its semigroup is a Feller semigroup.
Theorem 1.6. The transition semigroup $\left(Q_{t}\right)_{t \geq 0}$ defined by Equation(1.22) is a Feller semigroup.

Proof. For $\lambda \geq 0$ and $x \geq 0$ set $e_{\lambda}(x)=e^{-\lambda x}$. Denote with $\mathcal{D}_{0}$ the linear span of $e_{\lambda}: \lambda \geq 0$. By Proposition 1.4 the operator $Q_{t}$ preserves $\mathcal{D}_{0}$ for every $t \geq 0$. By the continuity of $t \mapsto v_{t}(\lambda)$ it is easy to show that $t \mapsto Q_{t} e_{\lambda}(x)$ is continuous for $\lambda \geq 0$ and $x \geq 0$. This implies that $t \mapsto Q_{t} f(x)$ is continuous for every $f \in \mathcal{D}_{0}$ and $x \geq 0$.
Let $C_{0}([0, \infty))$ the space of continuous functions on $[0, \infty)$ vanishing at $\infty$. By the StoneWeierstrass theorem the set $\mathcal{D}_{0}$ is uniformly dense in $C_{0}([0, \infty))$. Then each operator $Q_{t}$ preserves $C_{0}([0, \infty))$ and $t \mapsto Q_{t} f(x)$ is continuous for all $x \geq 0$ and $f \in C_{0}([0, \infty))$. This gives the Feller property of the semigroup $\left(Q_{t}\right)_{t \geq 0}$.

At this point we have all the ingredients to define a CB-process:
Definition 1.9. A Markov process on $[0, \infty)$ is called a continuous state Branching process (CB-Process) with branching mechanism $\varphi$ if it has transition semigroup $\left(Q_{t}\right)_{t \geq 0}$ defined by (1.22). Moreover the family of functions $\left(v_{t}\right)_{t \geq 0}$ defined by Equation (1.18) is called the cumulant semigroup of the CB-process.

It is easy to see that $\left(Q_{t}\right)_{t \geq 0}$ satisfies the branching property:

$$
Q_{t}\left(x_{1}+x_{2}, \cdot\right)=Q_{t}\left(x_{1}, \cdot\right) * Q_{t}\left(x_{2}, \cdot\right)
$$

To end this section we state a proposition which states an important property of CBprocesses:

Proposition 1.5. Suppose that $\left\{\left(X_{1}(t), \mathcal{F}_{t}^{1}\right): t \geq 0\right\}$ and $\left\{\left(X_{2}(t), \mathcal{F}_{t}^{2}\right): t \geq 0\right\}$ are two independent CB-processes with branching mechanism $\varphi$ and $\mathcal{F}_{t}^{1}, \mathcal{F}_{t}^{2}$ their associated $\sigma$ algebras. Let $X(t)=X_{1}(t)+X_{2}(t)$ and $\mathcal{F}_{t}=\sigma\left(\mathcal{F}_{t}^{1} \cup \mathcal{F}_{t}^{2}\right)$. Then the process $\left\{\left(X(t), \mathcal{F}_{t}\right)\right.$ : $t \geq 0\}$ is also a CB-process with branching mechanism $\varphi$.

Proof. Let $t \geq r \geq 0$ and $F_{r}^{i}$ be $\mathcal{F}_{r}^{i}$-measurable random variables for $i=1,2$. For any $\lambda \geq 0$ we have:

$$
\begin{aligned}
\mathbb{P}\left(F_{1} F_{2} e^{-\lambda X(t)}\right. & =\mathbb{P}\left(F_{1} e^{-\lambda X_{1}(t)}\right) \mathbb{P}\left(F_{2} e^{-\lambda X_{2}(t)}\right) \\
& =\mathbb{P}\left(F_{1} e^{-X_{1}(r) v_{t-r}(\lambda)}\right) \mathbb{P}\left(F_{2} e^{-X_{2}(r) v_{t-r}(\lambda)}\right. \\
& =\mathbb{P}\left(F_{1} F_{2} e^{-X(r) v_{t-r}(\lambda)}\right)
\end{aligned}
$$

A monotone class argument shows that:

$$
\mathbb{P}\left(F e^{-\lambda X(t)}\right)=\mathbb{P}\left(F e^{-X(r) v_{t-r}(\lambda)}\right)
$$

For any bounded $\mathcal{F}_{r}$-measurable random variable $F$. Then $\left\{\left(X(t), \mathcal{F}_{t}\right): t \geq 0\right\}$ is also a CB-process with transition semigroup $\left(Q_{t}\right)_{t \geq 0}$.

### 1.4 Construction of a CBI processe

In this section we present the construction of a CBI process starting from a Galton-Watson process with immigration; we follow [20].

### 1.4.1 Galton-Watson process with immigration

Let $\{p(j): j \in \mathbb{N}\}$ and $\{q(j): j \in \mathbb{N}\}$ be probability distribution on the space of positive integers $\mathbb{N}$ with generating functions $g$ and $h$ respectively. Suppose that $\left\{\xi_{n, i}: n, i=\right.$ $1,2, \ldots\}$ is a family of $\mathbb{N}$-valued random variables with distribution $\{p(j): j \in \mathbb{N}\}$ and $\left\{\eta_{n}: n=1,2, \ldots\right\}$ is a family on $\mathbb{N}$-valued random variables with distribution $\{q(j): j \in$ $\mathbb{N}\}$. Assume moreover that the two families are mutually independent. Given another random variable $Y(0)$ independent of $\left\{\xi_{n, i}\right\}$ and $\left\{\eta_{n}\right\}$ we define inductively:

$$
\begin{equation*}
Y(n)=\sum_{i=1}^{Y(n-1)} \xi_{n, i}+\eta_{n} \tag{1.23}
\end{equation*}
$$

This is clearly a generalization of Equation (1.4). For $i \in \mathbb{N}$ let $\{Q(i, j): j \in \mathbb{N}\}$ denote the $i$-fold convolution of $\{p(j): j \in \mathbb{N}\}$ by:

$$
P(i, j)=(Q(i, j) * q)(j)=\left(p^{* i} * q\right)(j)
$$

Now for $n \geq 1$ and $\left\{i_{0}, \ldots, i_{n-1}=i, j\right\}$ with $i_{j} \in \mathbb{N}$ we have:

$$
\begin{aligned}
P\left(Y(n)=j \mid Y(0)=i_{0}, \ldots, Y(n-1)=i_{n-1}\right) & =P\left(\sum_{k=1}^{Y(n-1)} \xi_{n, k}+\eta_{n}=j \mid Y(n-1)=i\right) \\
& =P\left(\sum_{k=1}^{Y(n-1)} \xi_{n, k}+\eta_{n}=j\right) .
\end{aligned}
$$

By the above computation $\{Y(n): n \in \mathbb{N}\}$ is a Markov chain with one step transition matrix given by: $P=\{P(i, j): i, j \in \mathbb{N}\}$.
Remark 1.1. The random variable $Y(n)$ can be thougth as the number of individuals in generation $n$ of a population system with immigration; after one unit of time each of the $Y(n)$ individuals splits independently of others into a random number of offspring according to the distribution $\{p(j): j \in \mathbb{N}\}$ and a random number of immigrants are added to the system according to the distribution $\{q(j): j \in \mathbb{N}\}$.

It is easy to see that:

$$
\begin{equation*}
\sum_{j=0}^{\infty} P(i, j) z^{j}=g(z)^{i} h(z) \quad|z| \leq 1 \tag{1.24}
\end{equation*}
$$

A Markov chain in $\mathbb{N}$ with one step transition matrix defined by (1.24) is called a GaltonWatson branching process with immigration (GWI-process) or a Bienaym $\tilde{A}^{*}$-GaltonWatson branching process with immigration (BGWI-process) with branching distribution given by $g$ and immigration distribution $h$. When $h \equiv 1$ this reduces to the previous GW process defined before. For any $n \geq 1$ the $n$-step transition matrix of the GWI-process is just the $n$-fold product $P^{n}=\left\{P^{n}(i, j): i, j \in \mathbb{N}\right\}$.

Proposition 1.6. For any $n \geq 1$ and $i \in \mathbb{N}$ we have:

$$
\begin{equation*}
\sum_{j=0}^{\infty} P^{n}(i, j) z^{j}=g^{\circ n}(z)^{i} \prod_{j=1}^{n} h\left(g^{\circ(j-1)}(z)\right) \quad|z| \leq 1 \tag{1.25}
\end{equation*}
$$

Proof. From (1.24) we have that the statement is true for $n=1$; suppose now that it holds for some $n \geq 1$, then:

$$
\begin{aligned}
\sum_{j=0}^{\infty} P^{n+1}(i, j) z^{j} & =\sum_{j=0}^{\infty} \sum_{k=0}^{\infty} P(i, k) P^{n}(k, j) z^{j} \\
& =\sum_{k=0}^{\infty} P(i, k) g^{\circ n}(z)^{k} \prod_{j=1}^{n} h\left(g^{\circ(j-1)}(z)\right) \\
& =g\left(g^{\circ n}(z)\right)^{i} h\left(g^{\circ n}(z)\right) \prod_{j=1}^{n} h\left(g^{\circ(j-1)}(z)\right) \\
& =g^{\circ(n+1)}(z)^{i} \prod_{j=1}^{n+1} h\left(g^{\circ(j-1)}(z)\right)
\end{aligned}
$$

Suppose now to have a sequence of GWI-processes $\left\{Y_{k}(n): n \geq 0\right\}$ with branching distribution given by the probability generating function $g_{k}$ and immigration distribution given by the probability generating function $h_{k}$. Let:

$$
Z_{k}(n):=\frac{Y_{k}(n)}{k} \quad k \in \mathbb{N}, n \in \mathbb{N}
$$

Then $\left\{Z_{k}(n): n \geq 0\right\}$ is a Markov chain with state space $E_{k}=\left\{0, k^{-1}, 2 k^{-1}, \ldots\right\}$ and $n$-step transition probability $P_{k}^{n}(x, d y)$ determined by:

$$
\begin{equation*}
\int_{E_{k}} e^{-\lambda y} P_{k}^{n}(x, d y)=g_{k}^{\circ n}\left(e^{-\lambda x}\right)^{k} x \prod_{j=1}^{n} h_{k}\left(g_{k}^{\circ(j-1)}\left(e^{-\lambda / k}\right)\right) \tag{1.26}
\end{equation*}
$$

Suppose now that $\left\{\gamma_{k}\right\}$ is a positive real sequence so that $\gamma_{k} \rightarrow \infty$ as $k \rightarrow \infty$. Let $\left\lfloor\gamma_{k} t\right\rfloor$ denote the integer part of $\gamma_{k} t$. In view of (1.26) given $Z_{k}(0)=x \in \mathbb{E}_{\boldsymbol{ך}}$ the random variable:

$$
Z_{k}\left(\left\lfloor\gamma_{k} t\right\rfloor\right)=\frac{Y_{k}\left(\left\lfloor\gamma_{k} t\right\rfloor\right)}{k}
$$

has distribution $P_{k}^{\left\lfloor\gamma_{k} t\right\rfloor}(x, \cdot)$ determined by:

$$
\begin{aligned}
\int_{E_{k}} e^{\lambda y} P_{k}^{\left\lfloor\gamma_{k} t\right\rfloor}(x, d y) & =g_{k}^{\left\lfloor\gamma_{k} t\right\rfloor}\left(e^{-\lambda / k}\right)^{k x} \prod_{j=1}^{\left\lfloor\gamma_{k} t\right\rfloor} h_{k}\left(g_{k}^{\circ(j-1)}\left(e^{-\lambda / k}\right)\right) \\
& =\exp \left\{x k \log \left[g_{k}^{\left\lfloor\gamma_{k} t\right\rfloor}\left(e^{-\lambda / k}\right)\right]\right\} \exp \left\{\sum_{j=1}^{\left\lfloor\gamma_{k} t\right\rfloor} \log \left[h_{k}\left(g_{k}^{\circ(j-1)}\right)\left(e^{-\lambda / k}\right)\right]\right\} \\
& =\exp \left\{-x v_{k}(t, \lambda)-\int_{0}^{\gamma_{k}^{-1}\left\lfloor\gamma_{k} t\right\rfloor} \psi_{k}\left(v_{k}(s, k)\right) d s\right\}
\end{aligned}
$$

where:

$$
v_{k}=-k \log \left(g_{k}^{\circ\left\llcorner\gamma_{k} t\right\rfloor}\left(e^{-\lambda / k}\right)\right)
$$

and

$$
\begin{equation*}
\psi_{k}(z)=-\gamma_{k} \log h_{k}\left(e^{-z / k}\right) \tag{1.27}
\end{equation*}
$$

As in Section 1.3 we can rewrite Equation (1.27) in term of a function $\tilde{\psi}(z)$ :

$$
\begin{equation*}
\psi_{k}(z)=-\gamma_{k} \log \left[1-\gamma_{k}^{-1} \tilde{\psi}_{k}(z)\right] \tag{1.28}
\end{equation*}
$$

where:

$$
\begin{equation*}
\tilde{\psi}_{k}(z)=\gamma_{k}\left[1-h_{k}\left(e^{-z / k}\right)\right] . \tag{1.29}
\end{equation*}
$$

Lemma 1.3. Suppose that the sequence $\left\{\tilde{\psi}_{k}\right\}_{k}$ is uniformely bounded on each bounded interval. Then we have that $\lim _{k \rightarrow \infty}\left|\tilde{\psi}_{k}(z)-\psi_{k}(z)\right|=0$ uniformely on each bounded interval.

We set up now a condition that will be used in the following theorems:
Assumption 1.2. There is a function $\psi$ on $[0, \infty)$ such that $\tilde{\psi}_{k}(z) \rightarrow \psi(z)$ uniformly on $[0, a]$ for every $a \geq 0$ as $k \rightarrow \infty$.

Proposition 1.7. Suppose that Assumption 1.2 is satisfied. Then the limit function $\psi$ has representation:

$$
\begin{equation*}
\psi(z)=\beta z+\int_{(0, \infty)}\left(1-e^{-z u}\right) \nu(d u) \tag{1.30}
\end{equation*}
$$

where $\beta \geq 0$ is a positive constant and $\nu$ is a $\sigma$-finite measure on $(0, \infty)$ such that:

$$
\int_{(0, \infty)}(1 \wedge u) \nu(d u)<\infty
$$

Proof. It is well known that $\varphi$ has representation (1.30) if and only if $e^{-\psi}=L_{\mu}$ is the Laplace transform of an infinitely divisible distribution $\mu$ on $[0, \infty$ ) (see [21] Theorem 1.39). In view of (1.29) the function can be represented by a special form (1.30), so $e^{-\tilde{\psi}_{k}}=L_{\mu_{k}}$ is the Laplace transform of an infinitely divisible distribution $\mu_{k}$ on $[0, \infty)$. By the previous lemma and Assumption 1.2 we have that: $\psi_{k}(z) \rightarrow \psi(z)$ uniformly on $[0, a]$ for every $a \geq 0$ as $k \rightarrow \infty$. By Theorem 1.2 there is a probability distribution $\mu$ on $[0, \infty)$ such that $\mu=\lim _{k \rightarrow \infty} \mu_{k}$ weakly and $e^{-\psi}=L_{\mu}$ clearly $\mu$ is also infinitely divisible and hence $\psi$ has representation (1.30).

The following two theorems have a proof which is the complete analoguous to the one presented in the previous section so it is omitted.

Proposition 1.8. For any function $\psi$ with representation (1.30) there is a sequence $\left\{\tilde{\psi}_{k}\right\}_{k}$ in the form (1.29) satisfying Assumption 1.2.

Theorem 1.7. Suppose that $\varphi$ and $\psi$ are given by (1.16) and (1.30) respectively. For any $\lambda \geq 0$ let $t \mapsto v_{t}(\lambda)$ be the unique solution of (1.18). Then there is a Feller semigroup $\left(P_{t}\right)_{t \geq 0}$ on $[0, \infty)$ defined by:

$$
\begin{equation*}
\int_{[0, \infty)} e^{-\lambda y} P_{t}(x, d y)=\exp \left\{-x v_{t}(\lambda)-\int_{0}^{t} \psi\left(v_{s}(\lambda)\right) d s\right\} \tag{1.31}
\end{equation*}
$$

We have all the ingredients to define a CBI process:
Definition 1.10. A Markov process on $[0, \infty)$ with transition semigroup defined by (1.31) is called a Continuous-state Branching process with Immigration (CBI-Process) with branching mechanism $\varphi$ given by Equation (1.16) and defined on $[0, \infty)$ and immigration mechanism $\psi$ given by (1.30) defined again on $[0, \infty)$.

The following proposition states an additivity property of CBI processes:
Proposition 1.9. Suppose that $\left\{\left(Y_{1}(t), \mathcal{G}_{t}^{1}\right): t \geq 0\right\}$ and $\left\{\left(Y_{2}(t), \mathcal{G}_{t}^{2}\right): t \geq 0\right\}$ are two independent CBI processes with branching mechanism $\varphi$ and immigration rate $\psi_{1}$ and $\psi_{2}$. Let:

$$
\begin{aligned}
& Y(t)=Y_{1}(t)+Y_{2}(t) \\
& \mathcal{G}_{t}=\sigma\left(\mathcal{G}_{t}^{1} \cup \mathcal{G}_{t}^{2}\right)
\end{aligned}
$$

Then the process $\left\{\left(Y(t), \mathcal{G}_{t}\right): t \geq 0\right\}$ is a CBI-process with branching mechanism $\varphi$ and immigration rate $\psi=\psi_{1}+\psi_{2}$.

Proof. Let $t \geq r \geq 0$ and for $i=1,2$ let $F_{i}$ a bounded positive $\mathcal{G}_{t}^{i}$-measurable random variable. For any $\lambda \geq 0$ we have:

$$
\begin{aligned}
\mathbb{P}\left[F_{1} F_{2} e^{-\lambda Y(t)}\right] & =\mathbb{P}\left[F_{1} e^{-\lambda Y_{1}(t)}\right] \mathbb{P}\left[F_{2} e^{-\lambda Y_{2}(t)}\right] \\
& =\mathbb{P}\left[F_{1} \exp \left\{-Y_{1}(t) v_{t-r}(\lambda)-\int_{0}^{t-r} \psi_{1}\left(v_{s}(\lambda)\right) d s\right\}\right] \\
& \cdot \mathbb{P}\left[F_{2} \exp \left\{-Y_{2}(t) v_{t-r}(\lambda)-\int_{0}^{t-r} \psi_{2}\left(v_{s}(\lambda)\right) d s\right\}\right] \\
& =\mathbb{P}\left[F_{1} F_{2} \exp \left\{-Y(t) v_{t-r}(\lambda)-\int_{0}^{t-r} \psi\left(v_{s}(\lambda)\right) d s\right\}\right]
\end{aligned}
$$

As in the proof of proposition 1.5 one can see that $\left\{\left(Y(t), \mathcal{G}_{t}\right): t \geq 0\right\}$ is a CBI process with branching mechanism $\varphi$ and immigration rate $\psi$.

To conclude this Section we give an additional definition of CBI processes which will be the base of the models that we are going to build in the following chapter.

Definition 1.11. A Markov process $Y$ with state space $\mathbb{R}^{+}$is called a CBI process characterized by branching mechanism $\varphi$ and immigration rate $\psi$ if its characteristic representation, for $p \geq 0$, is given by:

$$
\begin{equation*}
E_{y}\left[e^{-p Y(t)}\right]=\exp \left(-y \nu(t, p)-\int_{0}^{t} \Phi(\nu(s, p)) d s\right) \tag{1.32}
\end{equation*}
$$

where $E_{y}$ denotes the conditional expectation of with respect to the initial value $Y(0)=y$. The function $\nu: \mathbb{R}^{+} \times \mathbb{R}^{+} \rightarrow \mathbb{R}_{+}$satisfies the following differential equation:

$$
\left\{\begin{array}{l}
\frac{\partial \nu(t, p)}{\partial t}=-\psi(\nu(t, p))  \tag{1.33}\\
\nu(0, p)=p
\end{array}\right.
$$

where $\psi$ and $\Phi$ are functions of the variable $q \geq 0$ given by:

$$
\begin{aligned}
& \varphi(q)=a q+\frac{1}{2} \sigma^{2} q^{2}+\gamma \int_{0}^{\infty}\left(e^{-q u}-1+q u\right) \pi(d u) \\
& \psi(q)=a b q+\int_{0}^{\infty}\left(1-e^{-q u}\right) \nu(d u)
\end{aligned}
$$

with $a, b \in \mathbb{R}, \sigma, \gamma \geq 0$ and $\pi, \nu$ being two Lévy measures such that:

$$
\begin{aligned}
& \int_{0}^{\infty}\left(u \wedge u^{2}\right) \pi(d u)<\infty \\
& \int_{0}^{\infty}(1 \wedge u) \nu(d u)<\infty
\end{aligned}
$$

An important theorem was proved by Li (see [12] Theorem 3.1) and it guarantees the uniqueness od a solution to the SDE given by (??):

Theorem 1.8. There is a unique non-negative strong solution of the SDE:
$Y(t)=Y(0)+\int_{0}^{t} a(b-Y(s)) d s+\sigma \int_{0}^{t} \int_{0}^{Y(s)} W(d s, d u)+\gamma \int_{0}^{t} \int_{0}^{Y\left(s^{-}\right)} \int_{\mathbb{R}^{+}} z \tilde{N}(d s, d u, d z)$
Moreover the solution $\left\{Y_{t}: t \geq 0\right\}$ is a CBI process with branching mechanism:

$$
\psi(q)=a q+\frac{1}{2} \sigma^{2} q^{2}+\gamma \int_{0}^{\infty}\left(e^{-q \gamma z}-1+q \gamma z\right) \mu(d z)
$$

And immigration rate given by:

$$
\Phi(q)=a b q \quad q \geq 0
$$

### 1.5 Hawkes Processes

The self exciting point process, which is commonly known as the Hawkes process, it is a point process on the real line introduced by Hawkes (1971). The distinguishing feature of such processes is that they allow all past events to affect the intensity function at the current time. In this section we present the definition and some important features of Hawkes processes and then we provide the definition of Hawkes processes with exponential kernel that will be used in the following chapter.

### 1.5.1 Unmarked point processess

An unmarked random point process on the nonnegative real line $[0, \infty)$, where the nonnegative line is taken to represent time, is a random process whose realisations consists on a sequence of times $T_{1}, T_{2}, \ldots$ where $T_{n}$ is the $n$-th occurrence of an event. We make moreover the assumption of non-explosion:

$$
\begin{equation*}
T_{\infty}=\lim _{n \rightarrow \infty} T_{n}=\infty \tag{1.34}
\end{equation*}
$$

this means that almost surely we do not have an accumulation of events in finite time. Given a random point process $\left\{T_{n}\right\}_{n \in \mathbb{N}}$ the sequence of inter-events is defined as $\left\{S_{n}\right\}_{n \in \mathbb{N}}$ :

$$
\begin{equation*}
S_{n}=T_{n}-T_{n-1} . \tag{1.35}
\end{equation*}
$$

To represent a point process we can introduce the following:
Definition 1.12. Let $N:([0, \infty), \mathcal{B}[0, \infty)) \rightarrow \mathbb{N}$ we say that $N$ is the counting measure associated to an unmarked point process if for any $A \in \mathcal{B}([0, \infty)$ we have that:

$$
N(A)=\#\left\{i: T_{i} \in A\right\} .
$$

Clearly if $A=[a, b)$ then we have that:

$$
N(A)=\sum_{i=1}^{\infty} \mathbf{1}_{\left\{a \leq T_{i}<b\right\}}
$$

In particular we set:

$$
N_{t}=N([0, t))=\sum_{i=1}^{\infty} \mathbf{1}_{\left\{0 \leq T_{i}<t\right\}} .
$$

The point process defined above may be represented equivalently via its associated counting process $\left\{N_{t}\right\}_{t \geq 0}$ where:

$$
\begin{equation*}
N_{t}=n \text { if } t \in\left[T_{n}, T_{n}+1\right) \tag{1.36}
\end{equation*}
$$

In particular the random variable $N_{t}$ counts the number of events up to time $t$ and the non-explosion condition becomes $N_{t}<\infty$. Both the processes $\left\{T_{n}\right\}_{n \in \mathbb{N}}$ and $\left\{N_{t}\right\}_{t \geq 0}$ are the defined on some probability space $(\Omega, \mathcal{F}, P)$ with a filtration $\{\mathcal{F}\}_{t \geq 0}$ for which $N$ is $\left\{\mathcal{F}_{t}\right\}_{t \geq 0}$ adapted. One important example of a random point process is:
Definition 1.13. A point process $\left\{N_{t}\right\}_{t \geq 0}$ is called a Poisson point process if:

1. $N_{0}=0$
2. $N_{t}$ is a process with independent increments i.e. $\left(N_{t}-N_{s}\right) \perp \mathcal{F}_{s} \forall t>s$.
3. $\left(N_{t}-N_{s}\right)$ is a Poisson random variable with parameter $\Lambda_{s, t}=\lambda(t-s)=\int_{s}^{t} \lambda(u) d u$

Usually one assumes that $\Lambda_{s, t}=\int_{s}^{t} \lambda_{u} d u$ for some $\lambda_{t}$ deterministic function called intensity of the Poisson Point Process. In particular if $\left\{\mathcal{F}_{t}\right\}_{t \geq 0}$ is the filtration generated by $N_{t}$ i.e. $\mathcal{F}_{t}=\sigma\left\{N_{s}: s \leq t\right\}$ and $\lambda_{t} \equiv 1$ then $N_{t}$ is called standard Poisson process. One can also prove that if $\lambda_{t}=\lambda$ then the sequence of inter-events $\left\{S_{n}\right\}_{n \in \mathbb{N}}$ is made of i.i.d. exponential random variables with parameter $\lambda$. We now focus our attention to the case on which the intensity is $\lambda$ constant. By definition we have that:

$$
\begin{aligned}
& P\left(\left(N_{t}-N_{s}\right)=0\right)=e^{-\lambda(t-s)} \frac{\lambda^{0}(t-s)^{0}}{0!}=e^{\lambda(t-s)}=1-\lambda \cdot \Delta+o(\Delta) \\
& P\left(\left(N_{t}-N_{s}\right)=1\right)=e^{-\lambda(t-s)} \frac{\lambda(t-s)}{1!}=(t-s) e^{\lambda(t-s)}=\lambda \cdot \Delta+o(\Delta) \\
& P\left(\left(N_{t}-N_{s}\right) \geq 2\right)=\sum_{k \geq 2} e^{-\lambda(t-s)} \frac{\lambda^{k}(t-s)^{k}}{k!}=o(\Delta)
\end{aligned}
$$

where $\Delta=t-s$.

Remark 1.2. The above characterization for a Poisson process is parallel to the one of a Wiener process:

- both are processes with independent increment; the increments of a Wiener process are normally distributed while those of a Poisson process are Poisson distributed.
- The Wiener process is a basic building block for processes with continuous trajectories, the Poisson process is a basic building block for processes with jumping trajectories.
- the Wiener process is itself a martingale while the Poisson process is not, nevertheless it becomes a martingale if one subtracts from $N_{t}$ its Compensator.

Lemma 1.4. Given $(\Omega, \mathcal{F}, P)$ a probability space with filtration $\mathcal{F}_{t},\left\{N_{t}\right\}_{t \geq 0}$ Poisson process with intensity $\lambda_{s} \in L^{1}(0, s)$ which is $\mathcal{F}_{t}$ adapted we have that the process:

$$
\begin{equation*}
M_{t}=N_{t}-\int_{0}^{t} \lambda_{s} d s \tag{1.37}
\end{equation*}
$$

is an $\mathbb{P}$ martingale
Proof. We have to check that $E\left(M_{t} \mid \mathcal{F}_{s}\right)=M_{s} \forall t>s$, in particular we have:

$$
\begin{aligned}
E\left(M_{t}-M_{s} \mid \mathcal{F}_{s}\right) & =E\left(N_{t}-N_{s}-\int_{s}^{t} \lambda_{u} d u \mid \mathcal{F}_{s}\right)=E\left(N_{t}-N_{s} \mid \mathcal{F}_{s}\right)-\int_{s}^{t} \lambda_{u} d u \\
& =E\left(N_{t}-N_{s}\right)-\int_{s}^{t} \lambda_{u} d u=0
\end{aligned}
$$

Where we used that $\left(N_{t}-N_{s}\right) \perp \mathcal{F}_{s}$ and $\left(N_{t}-N_{s}\right) \sim \operatorname{Pois}\left(\Lambda_{s, t}\right)$ and in particular $\Lambda_{s, t}=\int_{s}^{t} \lambda_{u} d u$

### 1.5.2 Marked point process

A Marked point process is a point process with a random variable or a random vector attached to each point; each of the times $T_{i}$ has a mark $M_{i}$ associated with it and a possible realization of a marked point process is given by a sequence:

$$
\left(T_{1}, M_{1}\right), \ldots,\left(T_{i}, M_{i}\right), \ldots
$$

with $T_{1}<T_{2}, \ldots$ and $M_{i} \in \mathcal{M}$ the space of marks (we usually work with non negative marks). The counting process associated is the process $\left\{N_{t}\right\}_{t \geq 0}$ which, as before, counts the number of occurrence of the sequence of the $T_{i}$ in the interval $[0, t)$.

Remark 1.3. The generalization of unmarked point processes to marked point processes subsumes several important point processes. For example, a marked point process can be used to define a point process with multiple occurrences; the marks in this case would give the number of occurrences at each point event. A marked point process may also be used to define a multi-type point process with the marks identifying the type of a point event, i.e. $M=\{1,2, \ldots, k\}$ for a multi-type point process with $k$ types of point event.

### 1.5.3 Hawkes processes

We start defining the univariate Hawkes process recalling that $\left\{\mathcal{F}_{t}\right\}_{t \geq 0}$ is $\sigma$-algebra generated containing the complete history of the process:

Definition 1.14. The univariate Hawkes process $N$ with intensity $\lambda\left(\cdot \mid \mathcal{F}_{t}\right)$ is defined for all $t \geq 0$ and $h \rightarrow 0^{+}$by:

$$
\begin{align*}
& P\left(N[t, t+h)=1 \mid \mathcal{F}_{t}\right)=\lambda\left(t \mid \mathcal{F}_{t}\right) h+o(h)  \tag{1.38}\\
& P\left(N[t, t+h)>1 \mid \mathcal{F}_{t}\right)=o(h)
\end{align*}
$$

where the complete intensity is defined by:

$$
\begin{align*}
\lambda\left(t \mid \mathcal{F}_{t}\right) & =\tau+\int_{[0, t)} \omega(t-u) N(d u) \\
& =\tau+\sum_{j: t_{j}<t} \omega\left(t-t_{j}\right) \tag{1.39}
\end{align*}
$$

with $\operatorname{\tau in} \mathbb{R}, \tau>0$ and $\omega(s) \geq 0$
The complete intensity is a stochastic process and can be tougth as a Shot Noise process (see [10] p.74) where all the past point events can contribute to the current value of the complete intensity.
The self-exciting nature of the Hawkes process arises via the integral in Equation (1.39). The contribution from a point event at time $t_{i}<t$ to the complete intensity at time $t$ is $\omega\left(t-t_{i}\right)$, and all points before time $t$ contribute in such a way to the complete intensity at time $t$. The function $\omega(\cdot)$ controls the effect that past point events have on the intensity and it is often assumed to be a monotonically decreasing function so that the latest point events have the greatest influence on the current value of the intensity.
For a monotonically decreasing $\omega(\cdot)$, the intensity will increase immediately after a point event and as time passes the effect from the point event dies off. As a result, the risk of further point events occurring increases immediately following a point event and this increased risk dies off as time passes.
A Markovian decay function is the exponential decay function which has the form:

$$
\begin{equation*}
\omega_{e}(s)=\psi \exp (-\gamma s) \tag{1.40}
\end{equation*}
$$

Where $\psi \geq 0, \gamma>0$ and $\psi<\gamma$. An other popular example of of a decay function used in several application is the power-law decay function which has the form:

$$
\begin{equation*}
\omega_{d}(s)=\frac{\psi}{(s+\gamma)^{n+1}} \tag{1.41}
\end{equation*}
$$

Where $\psi \geq 0, \eta, \gamma>0$ and $\psi<\eta \gamma^{\eta}$.
To conclude this Section we add a slightly different definition of Hawkes processes that will be used to in the following chapter:

Definition 1.15. We have:

- A point process with associated counting process $\left\{N_{t}\right\}_{t \geq 0}$,
- $J(d s, d z)$ a Poisson random measure with intensity $\lambda(t)$ satisfying the SDE:

$$
\begin{align*}
\lambda(t) & =\lambda(0)-\beta \int_{0}^{t} \lambda(s) d s+\alpha \int_{0}^{t} \int_{0}^{\infty} z J(d s, d z) \\
& =\exp (-\beta t) \lambda(0)+\alpha \sum_{i=1}^{N_{t}} \exp \left[-\beta\left(t-t_{i}\right)\right] Z_{i} \tag{1.42}
\end{align*}
$$

with $\beta>0$ is the rate of exponential decay of the influence of previous jumps on the intensity level, $\alpha$ is the amplitude of the of the memory kernel, $T_{i}$ the jump times and $Z_{i}$ the jump sizes.

Then we define an Hawkes process with exponential kernel a process that can be written as:

$$
\begin{equation*}
Y(t)=Y(0)+\sum_{i=0}^{N_{t}} Z_{i}=Y(0)+\int_{0}^{t} \int_{0}^{\infty} z J(d s, d z) \tag{1.43}
\end{equation*}
$$

Moreover we assume that the jump sizes are distributed according to an exponential density with parameter $\delta$ (so that only positive jumps appear in Equations (1.42) and (1.43)) and we can write:

$$
\tilde{J}(d s, d z)=J(d s, d z)-\lambda(s) \mu(d z) d s
$$

and $\mu(d z)=\delta \exp (-\delta z) d z$. The last Equation denotes the compensated version of the Poisson measure $J(d s, d z)$. Finally, in order to guarantee the non explosiveness of of the Hawkes process we assume that $\beta-\frac{\alpha}{\delta}>0$.

## Chapter 2

## Hawkes and CBI to model forward prices

In this chapter the first part is devoted to a review of some notions martingale approach and arbitrage theory in the Black \& Scholes market; the main references for this part are the books [4] and [14]. In the second part I will present two models for forward prices, where the dynamic of the underlying factor is driven by a CBI process and by an Hawkes process, applied to the power marke; the main references for this part will be the articles [8], [15] and [3].

### 2.1 Martingale approach to arbitrage theory

### 2.1.1 The general case

We will start by considering the special case when one of the assets on the market is a risk free asset with zero rate of return. As the basic setup we thus consider a financial market consisting of $N$ given risky traded assets, and the asset price vector is as usual denoted by:

$$
S(t)=\left(\begin{array}{c}
S_{1}(t)  \tag{2.1}\\
\vdots \\
S_{n}(t)
\end{array}\right)
$$

We also assume that there exists a risk free asset with price process $S_{0}(t)$ which satisfies the following:

$$
\begin{equation*}
S_{0}(t)>0 \quad \text { P-a.s. } \forall t \geq 0 \tag{2.2}
\end{equation*}
$$

The main problem is to give condition for absence of arbitrage in this model and this are easily obtained moving to the "normalized" economy where we use $S_{0}$ as a numeraire. Thus instead of looking at the price vector process $S=\left[S_{0}, S_{1}, \ldots, S_{n}\right]$, we look to the relative price vector process $S(t) / S_{0}(t)$ where we have used $S_{0}$ as a numeraire price.

Definition 2.1. The normalized economy (also referred as "Z-economy") is defined by the price vector process $Z$ where:

$$
Z(t)=\frac{S(t)}{S_{0}(t)}
$$

i.e.

$$
\begin{equation*}
Z(t)=\left[Z_{0}(t), \ldots, Z_{n}(t)\right]=\left[1, \frac{S_{1}(t)}{S_{0}(t)}, \ldots, \frac{S_{n}(t)}{S_{0}(t)}\right] \tag{2.3}
\end{equation*}
$$

Note that at this point we have two price systems to keep track of: the $S$-system and the $Z$-system; the following definition clarifies the the relations between the two systems:

Definition 2.2. We define:

- A portfolio strategy is any adapted $N+1$-dimensional process:

$$
h(t)=\left[h_{0}(t), h_{1}(t), \ldots, h_{n}(t)\right]
$$

- The $S$-value process $V^{S}(t, h)$ corresponding to the portfolio $h$ is given by:

$$
\begin{equation*}
V^{S}(t, h)=\sum_{i=0}^{N} h_{i}(t) S_{i}(t) \tag{2.4}
\end{equation*}
$$

- The Z-value process $V^{Z}(t, h)$ corresponding to the portfolio $h$ is given by:

$$
\begin{equation*}
V^{Z}(t, h)=\sum_{i=0}^{N} h_{i}(t) Z_{i}(t) \tag{2.5}
\end{equation*}
$$

- A portfolio is said to be admissible (as a Z-portfolio) if there exist a non negative real number $\alpha$ such that:

$$
\begin{equation*}
\int_{0}^{t} h_{S}(u) d Z(u) \geq-\alpha \quad \text { for all } t \in[0, T] \tag{2.6}
\end{equation*}
$$

- An admissible portfolio is said to be $S$-self financing if:

$$
\begin{equation*}
d V^{S}(t, h)=\sum_{i=0}^{N} h_{i}(t) d S_{i}(t) \tag{2.7}
\end{equation*}
$$

- An admissible portfolio is said to be Z-self financing if:

$$
\begin{equation*}
d V^{Z}(t, h)=\sum_{i=0}^{N} h_{i}(t) d Z_{i}(t) \tag{2.8}
\end{equation*}
$$

We have the following important lemma:
Lemma 2.1. With assumption and notations of the previous definition we have that:

- A portfolio $h$ is $S$-self financing if and only if it is $Z$-self financing.
- The value process $V^{Z}$ and $V^{S}$ are connected by:

$$
V^{Z}(t, h)=\frac{1}{\left.S_{0}(t)\right)} \cdot V^{S}(t, h)
$$

- A claim $\mathcal{Y}$ is $S$-reachable if and only if the claim:

$$
\frac{\mathcal{Y}}{S_{0}(t)}
$$

is Z-reachable.

- The model is $S$ arbitrage free if and only if it is $Z$ arbitrage free.

Proof. We just prove the third item since the other two are obvious. We assume moreover that all processes have stochastic differentials driven by a finite number of Wiener processes and that the portfolio $h$ is $S$-self financing. We set $\beta=S_{0}$, then we have:

$$
\begin{aligned}
& Z=\beta^{-1} S \\
& V^{S}=h \cdot S \\
& V^{Z}=\beta^{-1} V^{S} \\
& d V^{S}=h \cdot d S
\end{aligned}
$$

And we want to prove that:

$$
d V^{Z}=h \cdot d Z
$$

Using the Ito formula on $Z=\beta^{-1} S$ we thus want to prove that:

$$
d V^{Z}=\beta^{-1} h \cdot d S+h \cdot S d B^{-1}+h \cdot d S d \beta^{-1}
$$

From $V^{Z}=\beta^{-1} V^{S}$ we have:

$$
\begin{aligned}
d V^{Z} & =\beta^{-1} d V^{S}+V^{S} d \beta^{-1}+d \beta^{-1} d V^{S} \\
& =\beta^{-1} h \cdot d S+h \cdot S d \beta^{-1}+d \beta^{-1} h \cdot d S
\end{aligned}
$$

where we used $d V^{S}=h \cdot d S$ and $V^{S}=h \cdot S$.
We formulate the first fundamental theorem concerning absence of arbitrage:
Theorem 2.1. Consider the market model $S_{0}, S_{1}, \ldots, S_{N}$ where we assume that $S_{0}(t)>0$ $P$-a.s. for all $t \geq 0$. Assume furthermore that $S_{0}, S_{1}, \ldots, S_{N}$ are locally bounded. The model is arbitrage free if and only if there exist a martingale measure $Q \sim P$ such that the processes:

$$
Z_{0}, Z_{1}, \ldots, Z_{n}
$$

are local martingales under $Q$.

### 2.1.2 Completeness

In this section we assume absence of arbitrage i.e. that there exist a local martingale measure. We now turn to the possibility of replicating a given contingent claim in terms of a portfolio based on the underlying assets. We introduce an important lemma:

Lemma 2.2. Consider a given T-claim $X$. Fix a martingale measure $Q$ and assume that the normalized claim $X / S_{0}(t)$ is integrable. If the $Q$-martingale $M$ defined by

$$
\begin{equation*}
M(t)=E^{Q}\left[\left.\frac{X}{S_{0}(T)} \right\rvert\, \mathcal{F}_{t}\right] \tag{2.9}
\end{equation*}
$$

admits an integral representation of the form:

$$
\begin{equation*}
M(t)=x+\sum_{i=1}^{N} \int_{0}^{t} h_{i}(s) d Z_{i}(s) \tag{2.10}
\end{equation*}
$$

then $X$ can be hedged in the $S$-economy. Furthermore the replicating portfolio $\left(h_{0}, h_{1}, \ldots, h_{N}\right)$ is given by equation (2.10) for the part $\left(h_{1}, \ldots, h_{N}\right)$ whereas:

$$
\begin{equation*}
h_{0}(t)=M(t)-\sum_{i=1}^{N} h_{i}(t) Z_{i}(t) \tag{2.11}
\end{equation*}
$$

Proof. We want to hedge $X$ in the $S$-economy i.e. we want to hedge $X / S_{0}(t)$ in the $Z$ economy. In terms of normalized prices we are looking for a process $\left(h_{0}, h_{1}, \ldots, h_{N}\right)$ such that:

$$
\begin{equation*}
V^{Z}(T, h)=\frac{X}{S_{0}(T)} \quad \text { P-a.s. } \tag{2.12}
\end{equation*}
$$

and

$$
\begin{equation*}
d V^{Z}=\sum_{i=1}^{N} h_{i} d Z_{i} \tag{2.13}
\end{equation*}
$$

where the normalized value process is given by:

$$
\begin{equation*}
V^{Z}(t, h)=h_{0}(t) \cdot 1+\sum_{i=1}^{N} h_{i}(t) Z_{i}(t) \tag{2.14}
\end{equation*}
$$

A reasonable guess is that $M=V^{Z}$ so let $M$ be defined by (2.9) and set $\left(h_{1}, \ldots, h_{N}\right)$ by (2.10) and $h_{0}$ by:

$$
h_{0}(t)=M(t)-\sum_{i=1}^{N} h_{i}(t) Z_{i}(t)
$$

Now from (2.14) we obviously have $M=V^{Z}$ and from (2.10) we get:

$$
d V^{Z}=d M=\sum_{i=1}^{N} h_{i} Z_{i}(t)
$$

which shows that the portfolio is self financing. Furthermore we have:

$$
V^{Z}(T, h)=M(T)=E^{Q}\left[\left.\frac{X}{S_{0}(T)} \right\rvert\, \mathcal{F}_{T}\right]=\frac{X}{S_{0}(T)}
$$

which shows that $X$ is replicated by $h$.

We thus see that, modulo some integrability considerations, completeness is equivalent to the existence of a martingale representation theorem for the discounted price process. Now we are ready to set an important theorem which is the base of the proof of the second fundamental theorem of asset pricing:

Theorem 2.2. Let $\mathcal{M}$ denote the convex set of equivalent martingale measures. Then for any fixed $Q \in \mathcal{M}$ the following are equivalent:

- Every $Q$ local martingale $M$ has dynamics of the form:

$$
d M(t)=\sum_{i=1}^{N} h_{i}(t) d Z_{i}(t)
$$

- $Q$ is an extremal point of $\mathcal{M}$

Then we have:
Theorem 2.3 (Second fundamental theorem of asset pricing). Assume that the market is arbitrage free and consider a fixed numeraire asset $S_{0}$. Then the market is complete if and only if the martingale measure corresponding to the numeraire $S_{0}$ is unique.

Proof. If the martingale measure is unique then $\mathcal{M}$ is a singleton $\mathcal{M}=Q$ so $Q$ is trivially an extremal point of $\mathcal{M}$ and thus for the previous theorem (Jacod) we have a stochastic integral representation of every $Q$ martingale and then it follows from lemma 2.2 that the model is complete. The other implication is trivial.

### 2.1.3 Martingale pricing

We turn to the pricing problem for contingent claims. We thus consider the primary market $S_{0}, S_{1}, \ldots, S_{N}$ and we fix a $T$-claim $X$. Our task is to determine a reasonable price process $\Pi(t, X)$ assuming that the market is arbitrage free. There are two possible approaches:

- The derivative should be priced in a way that it is consistent with the price of the underlying assets. More precisely we should demand that the extended market $\Pi(t, X), S_{0}, S_{1}, \ldots, S_{N}$ if free of arbitrage opportunities.
- If the claim is attainable with hedging portfolio $h$ then the only reasonable price is given by $\Pi(t, X)=V(t, h)$.

In the first approach the aim is to find a martingale measure $Q$ for the extended market $\Pi(\cdot, X), S_{0}, S_{1}, \ldots, S_{N}$. Assuming that such a measure $Q$ does exist, and assuming enough integrability conditions, by definition of martingale measure we have that:

$$
\begin{equation*}
\frac{\Pi(t, X)}{S_{0}(t)}=E^{Q}\left[\left.\frac{\Pi(T ; X)}{S_{0}(T)} \right\rvert\, \mathcal{F}_{t}\right]=E^{Q}\left[\left.\frac{X}{S_{0}(T)} \right\rvert\, \mathcal{F}_{t}\right] \tag{2.15}
\end{equation*}
$$

We thus have the following:

Theorem 2.4 (General pricing formula). The arbitrge free price process for the $T$-claim $X$ is given by:

$$
\begin{equation*}
\Pi(t, X)=S_{0}(t) E^{Q}\left[\left.\frac{X}{S_{0}(T)} \right\rvert\, \mathcal{F}_{t}\right] \tag{2.16}
\end{equation*}
$$

Where $Q$ is the (not necessarily unique) martingale measure for the a-priori given market $S_{0}, S_{1}, \ldots, S_{N}$ with numeraire $S_{0}$.

If we assume that $S_{0}$ is the money account:

$$
\begin{equation*}
S_{0}(t)=S_{0}(0) \cdot e^{t_{0}^{t} r(s) d s} \tag{2.17}
\end{equation*}
$$

where $r$ denotes the short rate, then equation (2.16) reduces to the "risk-neutral valutation formula":

Theorem 2.5. Assuming the existence of a short rate, the pricing formula takes the form:

$$
\begin{equation*}
\Pi(t, X)=E^{Q}\left[X e^{\int_{t}^{T} r(s) d s} \mid \mathcal{F}_{t}\right] \tag{2.18}
\end{equation*}
$$

Where $Q$ is the (not necessarily unique) martingale measure with money account as numeraire.

For the second approach to pricing let us assume that $X$ can be replicated by $h$. Since the holding of the derivative contract and the holding of the replicating portfolio are equivalent from a financial point of view, we see that the price of the derivative must be given by the formula:

$$
\begin{equation*}
\Pi(t, X)=V(t, h) \tag{2.19}
\end{equation*}
$$

One problem is what happen in case when $X$ can be replicated by two different portfolios and one would also know how this formula is connected to (2.16).
Defining $\Pi(t, X)$ by (2.19) we see that $\Pi(t, X) / S_{0}(t)=V^{Z}(t)$ and since, assuming enough integrability, $V^{Z}$ is a $Q$ martingale we see that also $\Pi(t, X) / S_{0}(t, X)$ is a $Q$-martingale. Thus we obtain the formula:

$$
\begin{equation*}
V(t, h)=S_{0}(t) E^{Q}\left[\left.\frac{X}{S_{0}(T)} \right\rvert\, \mathcal{F}_{t}\right] \tag{2.20}
\end{equation*}
$$

which will hold for any replicating portfolio and any martingale measure $Q$.

### 2.2 The mathematics of Martingale approach

### 2.2.1 Stochastic integral representation

Let us consider a fixed time interval $[0, T]$, a probability space $(\Omega, \mathcal{F}, P)$ with some filtration $\left\{\mathcal{F}_{t}\right\}_{t \geq 0}$ and an adapted vector Wiener process $W=\left(W_{1}, \ldots, W_{d}\right)^{T}$. Fix now a vector process $h=\left(h_{1}, \ldots, h_{d}\right)$ which is "integrable enough" and a real number $x_{0}$.If we define the process:

$$
\begin{equation*}
M(t)=x_{0}+\sum_{i=1}^{d} \int_{0}^{t} h_{i}(s) d W_{i}(s), \quad t \in[0, T] \tag{2.21}
\end{equation*}
$$

then we know that $M$ is a martingale. (Recall that under mild integrability condition every stochastic integral w.r.t a Wiener process is an $\mathcal{F}_{t}$ martingale). A natural and important question is whether the converse holds, i.e. if every $\mathcal{F}_{t}$-adapted martingale $M$ can be written in the form (2.21); if this is the case we say that $M$ admits a stochastic integral representation.
It is not hard to see that in the completely general case, there is no hope for a stochastic integral representation w.r.t. to $W$ for a general martingale. As a counterexample consider the case in which $d=1$ so $W$ is scalar, we consider a Poisson process $N$ and we assume that $\mathcal{F}_{t}$ contains all the information generated by $W$ and $N$ in the interval $[0, t]$. We know that:

$$
M(t)=N(t)-\lambda t
$$

is an $\mathcal{F}_{t}$-martingale. If we look at the trajectories they consist into straight lines with downward slope $\lambda$ interrupted at exponentially distributed points in time by positive jumps of unit size. From this it is obvious that $M$ can posses no stochastic integral representation of the form (2.21) since any such representation has continuous trajectories. It is clear from the previous example that we can only hope for a stochastic integral representation in the case when $\left\{\mathcal{F}_{t}\right\}_{t \geq 0}$ is the internal filtration generated by the Wiener process $W$ itself.
We state now:
Theorem 2.6. Let $W$ a d-dimensional Wiener process and let $X$ a stochastic variable such that:

- $X \in \mathcal{F}_{T}^{W}$
- $E[|X|]<\infty$

There exist a uniquely determined $\mathcal{F}_{t}^{W}$-adapted process $h_{1}, \ldots, h_{d}$ such that $X$ has representation:

$$
\begin{equation*}
X=E[X]+\sum_{i=1}^{d} \int_{0}^{T} h_{i}(s) d W_{i}(s) \tag{2.22}
\end{equation*}
$$

Proof. For simplicity of notations we consider the scalar case $d=1$. Recall the Geometric Brownian motion equation:

$$
\begin{aligned}
& d X_{t}=\sigma X_{t} d W_{t} \\
& X_{0}=1
\end{aligned}
$$

and its solution:

$$
\begin{equation*}
X_{t}=e^{-\frac{1}{2} \sigma^{2} t+\sigma W_{t}} \tag{2.23}
\end{equation*}
$$

The previous SDE in integral form becomes:

$$
\begin{equation*}
X_{t}=1+\int_{0}^{t} \sigma X_{s} d W_{s} \tag{2.24}
\end{equation*}
$$

Plugging now (2.23) into (2.24) we obtain that:

$$
e^{\sigma W_{t}}=e^{\frac{1}{2} \sigma^{2} t}+\sigma \int_{0}^{t} e^{-\frac{1}{2} \sigma^{2}(u-t)+\sigma W_{u}} d W_{u}
$$

Using the same argument we obtain for $s \geq t$ that:

$$
\begin{equation*}
e^{\sigma\left(W_{t}-W_{s}\right)}=e^{\frac{1}{2} \sigma^{2}(t-s)}+\sigma \int_{s}^{t} e^{-\frac{1}{2} \sigma^{2}(u-t+s)+\sigma W_{u}} d W_{u} \tag{2.25}
\end{equation*}
$$

Thus any stochastic variable of the type:

$$
Z=\exp \left\{\sigma\left(W_{t}-W_{s}\right)\right\}
$$

will have representation of the form:

$$
Z=E[Z]+\int_{0}^{T} h_{u} d W_{u}
$$

with $h \equiv 0$ outside $[s, t]$. From this we have easily that any random variable $Z$ of the form:

$$
\begin{equation*}
Z=\prod_{k=1}^{n} \exp \left\{\sigma_{k}\left(W_{t_{k}}-W_{t_{k-1}}\right)\right\} \tag{2.26}
\end{equation*}
$$

where $0 \leq t_{0} \leq \cdots \leq t_{n} \leq T$ has representation of the form:

$$
\begin{equation*}
Z=E[Z]+\int_{0}^{T} h_{u} d W_{u} \tag{2.27}
\end{equation*}
$$

It is now fairly straightforward to see that any variable of the form:

$$
\begin{equation*}
Z=\prod_{k=1}^{n} \exp \left\{i \sigma_{k}\left(W_{t_{k}}-W_{t_{k-1}}\right)\right\} \tag{2.28}
\end{equation*}
$$

where $i$ is the imaginary unit, has representation of the form (2.27). Using now some Fourier techniques we can prove that the set of variables of the form (2.28) is dense in $L^{2}\left(\mathcal{F}_{T}\right)$ and from this we can conclude that any variable in $L^{2}\left(\mathcal{F}_{T}\right)$ has a representation of the type (2.27).

From this theorem we easily obtain the:
Theorem 2.7 (Martingale representation theorem). Let $W$ a d-dimensional Wiener process and assume that the filtration $\left\{\mathcal{F}_{t}\right\}_{t \geq 0}$ is defined as:

$$
\mathcal{F}_{t}=\mathcal{F}_{t}^{W}=\sigma\left\{W_{s}: s \leq t\right\} \quad t \in[0, T]
$$

Let $M$ any $\mathcal{F}_{t}$ adapted martingale, then there exist a uniquely determined $\mathcal{F}_{t}$ adapted process $h_{1}, \ldots, h_{d}$ such that $M$ has the representation:

$$
\begin{equation*}
M(t)=M(0)+\sum_{i=1}^{d} \int_{0}^{t} h_{i}(s) d W_{i}(s) \tag{2.29}
\end{equation*}
$$

The proof follows immediately from theorem 2.6.

### 2.2.2 The Girsanov Theorem

In this section we discuss the effect that an absolutely continuous measure transformation will have upon a Wiener process.
Assume that our space $\left(\Omega, \mathcal{F},\left\{\mathcal{F}_{t}\right\}_{t \geq 0}, P\right)$ carries a $P$-Wiener process $W^{P}$ and that for some fixed $T$ we have changed to a new measure $Q$ on $\mathcal{F}_{T}$ by choosing a non-negative random variable $L_{T} \in \mathcal{F}_{T}$ and defining $Q$ by:

$$
d Q=l_{T} d P \quad \text { on } \mathcal{F}_{\mathcal{T}}
$$

This measure transformation will generate a likelihood process $\left\{L_{t}: t \geq 0\right\}$ defined by:

$$
L_{t}=\frac{d Q}{d P} \quad \text { on } \mathcal{F}_{\mathcal{T}}
$$

and we know that the previous process is a $P$-martingale.
Since $L$ is a $P$-martingale and since any stochastic integral with respect to $W$ is a martingale it is natural to define $L$ as the solution of the stochastic differential equation:

$$
\left\{\begin{array}{l}
d L_{t}=\varphi_{t} L_{t} d W_{t}^{P}  \tag{2.30}\\
L_{0}=1
\end{array}\right.
$$

With this procedure it seems that we can generate a large class of natural measure transformations from $P$ to $Q$ by following:

1. Choose an arbitrary process $W$.
2. Define a likelihood process $L$ by:

$$
\left\{\begin{array}{l}
d L_{t}=\varphi_{t} L_{t} d W_{t}^{P} \\
L_{0}=1
\end{array}\right.
$$

3. Define a new measure $Q$ setting:

$$
d Q=L_{t} d P
$$

on $\mathcal{F}_{t}$ for all $t \in[0, T]$.
Applying Ito formula we easily find that

$$
L_{t}=e^{\int_{0}^{t} \varphi_{s} d W_{s}^{P}-\frac{1}{2} \int_{0}^{t} \varphi_{s}^{2} d s}
$$

Thus $L$ is non-negative and under some integrability conditions on $\varphi$ (Novikov condition) we easily derive that $L$ is a martingale with $E^{P}\left[L_{t}\right]=1$.
We recall now an important theorem known as "Abstract Bayes formula", (for a complete proof we remind to [4], Appendix B, proposition B.41)
Theorem 2.8. Assume that $X$ is a random variable on $(\Omega, \mathcal{F}, P)$ and let $Q$ a probability measure with radon Nikodym derivative:

$$
\begin{equation*}
L=\frac{d Q}{d P} \quad \text { on } \mathcal{F}_{t} \tag{2.31}
\end{equation*}
$$

Assume that $X \in L^{1}(\Omega)$ and that $\mathcal{G}$ is a sigma algebra s.t. $\mathcal{F} \subseteq \mathcal{G}$ then

$$
\begin{equation*}
E^{Q}[X \mid \mathcal{G}]=\frac{E^{P}[L \cdot X \mid \mathcal{G}]}{E^{P}[L \mid \mathcal{G}]} \quad Q-\text { a.s. } \tag{2.32}
\end{equation*}
$$

Our aim is now to compute what are the dynamics of $W^{P}$ under $Q$. Recall that if $X$ has dynamics:

$$
d X_{t}=\mu_{t} d t+\sigma_{t} d W_{t}^{P}
$$

and we have that:

$$
\begin{align*}
& E^{P}\left[d X_{t} \mid \mathcal{F}_{t}\right]=\mu_{t} d t \\
& E^{P}\left[\left(d X_{t}\right)^{2} \mid \mathcal{F}_{t}\right]=\sigma_{t}^{2} d t \tag{2.33}
\end{align*}
$$

with the informal interpretation $d X_{t}=X_{t+d t}-X_{t}$.
Let now $X=W^{P}$ (i.e. $\mu=0$ and $\sigma=1$ ), using theorem 2.8 and the fact that $L$ is a $P$-martingale we obtain

$$
\begin{aligned}
E^{Q}\left[d X_{t} \mathcal{F}_{t}\right] & =\frac{E^{P}\left[L_{t+d t} d X_{t} \mathcal{F}_{t}\right]}{E^{P}\left[L_{t+d t} \mathcal{F}_{t}\right]} \\
& =\frac{E^{P}\left[L_{t+d t} d X_{t} \mathcal{F}_{t}\right]}{L_{t}} \\
& =\frac{E^{P}\left[L_{t} d X_{t} \mathcal{F}_{t}\right]}{L_{t}}+\frac{E^{P}\left[d L_{t} d X_{t} \mathcal{F}_{t}\right]}{L_{t}}
\end{aligned}
$$

Since $L$ is adapted and $X$ has zero drift under $P$ we have that:

$$
\frac{E^{P}\left[L_{t} d X_{t} \mathcal{F}_{t}\right]}{L_{t}}=L_{t} \cdot \frac{E^{P}\left[d X_{t} \mathcal{F}_{t}\right]}{L_{t}}=0 \cdot d t
$$

Furthermore we have that:

$$
d L_{t} d X_{t}=L_{t} \varphi_{t} d W_{t}^{P}\left(0 \cdot d t+1 \cdot d W_{t}^{P}\right)=L_{t} \varphi_{t}\left(d W_{t}^{P}\right)^{2}
$$

and hence since $L_{t} \varphi_{t} \in \mathcal{F}_{t}$ :

$$
\frac{E^{P}\left[d L_{t} d X_{t} \mid \mathcal{F}_{t}\right]}{L_{t}}=\varphi_{t} d t
$$

finally using that $d X_{t}^{2}=d t$ we can compute the quadratic variation of $X$ under $Q$ :

$$
\begin{aligned}
E^{Q}\left[\left(d X_{t}\right)^{2} \mid \mathcal{F}_{t}\right] & =\frac{E^{P}\left[L_{t+d t}\left(d X_{t}\right)^{2} \mid \mathcal{F}_{t}\right]}{L_{t}} \\
& =\frac{E^{P}\left[L_{t+d t} d t \mid \mathcal{F}_{t}\right]}{L_{t}} \\
& =\frac{E^{P}\left[L_{t+d t} \mid \mathcal{F}_{t}\right]}{L_{t}} d t \\
& =d t
\end{aligned}
$$

We found that the process $X$ that was, under $P$, a standard Wiener process with unit diffusion term and zero drift, under $Q$ defined above the drift process has changed from zero to $\varphi$ and the diffusion remains the same. In other words we have that:

$$
d W_{t}^{P}=\varphi_{t} d t+d W_{t}^{Q}
$$

Theorem 2.9 (Girsanov Theorem). Let $W^{P}$ a d-dimensional P-Wiener process defined on a probability space $\left(\Omega, \mathcal{F},\left\{\mathcal{F}_{t}\right\}_{t \geq 0}, P\right)$ and let $\varphi$ be any d-dimensional adapted column vector process. Choose a fixed $T$ and define the process $L$ on $[0, T]$ by:

$$
\left\{\begin{array}{l}
d L_{t}=\varphi_{t}^{\star} L_{t} d W_{t}^{P}  \tag{2.34}\\
L_{0}=1
\end{array}\right.
$$

i.e.

$$
\begin{equation*}
L_{t}=e^{\int_{0}^{t} \varphi_{s}^{\star} d W_{s}^{P}-\frac{1}{2} \int_{0}^{t}\left\|\varphi_{s}\right\|^{2} d s} \tag{2.35}
\end{equation*}
$$

Assume that:

$$
\begin{equation*}
E^{P}\left[L_{T}\right]=1 \tag{2.36}
\end{equation*}
$$

and define a new probability measure $Q$ on $\mathcal{F}_{T}$ by:

$$
\begin{equation*}
\frac{d Q}{d P} \text { on } \mathcal{F}_{T} \tag{2.37}
\end{equation*}
$$

then

$$
\begin{equation*}
d W^{P}=\varphi_{t} d t+d W_{t}^{Q} \tag{2.38}
\end{equation*}
$$

where $W^{Q}$ is a $Q$-Wiener process.
Observation 2.1. In the previous theorem we adopted the following notation:

- we denote with $\varphi^{\star}$ the transpose of the vector $\varphi$ and we can rewrite equation (2.34) and (2.35) by components as:

$$
\begin{aligned}
& d L(t)=L(t) \sum_{i=1}^{d} \varphi_{i}(t) d W_{i}^{P}(t) \\
& L(t)=\exp \left\{\sum_{i=1}^{d} \int_{0}^{t} \varphi_{i}(s) d W_{i}^{P}(s)-\frac{1}{2} \int_{0}^{t} \sum_{i=1}^{d} \varphi_{i}^{2}(s) d s\right\}
\end{aligned}
$$

- We often refer to the process $\varphi$ as the Girsanov Kernel of the measure transformation

Definition 2.3. For any Wiener process $W$ and any kernel process $\varphi$ the Doleans-Dade exponential process $\mathcal{E}$ is defined by:

$$
\begin{equation*}
\mathcal{E}\left(\varphi^{\star} W\right)(t)=\exp \left\{\int_{0}^{t} \varphi^{\star}(s) d W^{P}(s)-\frac{1}{2} \int_{0}^{t}\|\varphi\|^{2}(s) d s\right\} \tag{2.39}
\end{equation*}
$$

Observe that in Girsanov theorem we have to assume an ad-hoc condition thet $\varphi$ is such that $E^{P}\left(L_{T}\right)=1$ or, in other words, that $L$ is a martingale. We can remove this hypothesis in case in which it is satisfied the following condition:

Lemma 2.3 (Novikov Condition). Assume that the Girsanov kernel $\varphi$ is such that:

$$
\begin{equation*}
E^{P}\left[e^{\frac{1}{2} \int_{0}^{T}\left\|\varphi_{t}\right\|^{2} d t}\right]<\infty \tag{2.40}
\end{equation*}
$$

Then $L$ is a martingale and in particular $E^{P}\left[L_{T}\right]=1$

### 2.3 Forward price modeling

In this section we are going to introduce two alternative models for the forward prices and we are going to test them against electricity market data. We assume that the spot price process evolves according to the following dynamics:

$$
\begin{equation*}
S(t)=\alpha(t)+X(t) \tag{2.41}
\end{equation*}
$$

where $\alpha(t)$ is a function representing the seasonality and the process $\{X(t): t \geq 0\}$ is a superposition of the factors $Y_{i}$; we can rewrite the previous equation as:

$$
\begin{equation*}
S(t)=\alpha(t)+\sum_{i=1}^{n} Y_{i}(t) \tag{2.42}
\end{equation*}
$$

### 2.4 The model based on CBI

In this subsection we will focus to the case in which the underlying factors follows a dynamic which is given by the stochastic differential equation:

$$
\begin{align*}
Y_{i}(t) & =Y_{i}(0)-\int_{0}^{t} \alpha_{i}\left(b_{i}-Y_{i}(s)\right) d s+\sigma_{i} \int_{0}^{t} \int_{0}^{Y_{i}(s)} W_{i}(d s, d u)  \tag{2.43}\\
& +\gamma_{i} \int_{0}^{t} \int_{0}^{Y_{i}(s-)} \int_{\mathbb{R}^{+}} x \tilde{N}(d s, d u, d z)
\end{align*}
$$

where $\alpha_{i}, \sigma_{i}, \gamma_{i} \in \mathbb{R}^{+}$are constant parameters for $i=1, \ldots, n$. Assuming that the factors $Y_{i}$ follows the dynamic given by equation (2.43) we obtain that:

Lemma 2.4. For any $T \geq 0$ and $\tau \leq T$ we have that:

$$
\begin{equation*}
E\left[Y_{i}(T) \mid \mathcal{F}_{\tau}\right]=b_{i}+\left(Y_{i}(\tau)-b_{i}\right) e^{-a_{i}(T-\tau)} \tag{2.44}
\end{equation*}
$$

Proof. First of all we observe that:

$$
\begin{aligned}
E\left(Y_{i}(T)\right) & =Y_{i}(0)+E\left[\int_{0}^{T} a_{i}\left(b_{i}-Y_{i}(s)\right) d s\right] \\
& =Y_{i}(0)+\int_{0}^{T} a_{i}\left(b_{i}-E\left(Y_{i}(s)\right)\right) d s
\end{aligned}
$$

set now $\varphi(t)=E\left(Y_{i}(t)\right)$, we can differentiate both sides to get:

$$
\frac{\partial}{\partial t} \varphi(t)=a_{i}\left(b_{i}-\varphi(t)\right)
$$

whose solution is:

$$
\begin{equation*}
E\left(Y_{i}(t)\right)=b_{i}+\left(Y_{i}(0)-b_{i}\right) e^{-a_{i} t} \tag{2.45}
\end{equation*}
$$

A direct application of Markov property allow us to conclude.

We discuss now the procedure of choosing a suitable risk neutral probability measure, in particular in the electricity market, the risk-neutral probability $\mathbb{Q}$ is often chosen by introducing a drift adjustment in the dynamics of an underlying asset. The following proposition tells us that after this change of measure the spot process remains in the same class (i.e. CBi processes with modified coefficients).

Proposition 2.1. Let $Y_{1}, \ldots, Y_{n}$ be independent CBI processes where for each $i=\{1, \ldots, n\}$ $Y_{i}$ is a CBI process under the historical probability measure $\mathbb{P}$. Assume that the filtration $\left\{\mathcal{F}_{t}\right\}_{t \geq 0}$ is generated by the random fields $W_{1}, \ldots, W_{n}$ and $\tilde{N}_{i}, \ldots, \tilde{N}_{n}$. For each $i$ fix $\eta_{i} \in \mathbb{R}$ and $\xi_{i} \in \mathbb{R}^{+}$and define:

$$
\begin{equation*}
U_{t}=\sum_{i=1}^{n} \eta_{i} \int_{0}^{t} \int_{0}^{Y_{i}(s)} W_{i}(d s, d u)+\sum_{i=1}^{n} \int_{0}^{t} \int_{0}^{Y_{i}(s-)} \int_{0}^{\infty}\left(e^{-\xi_{i} z}-1\right) \tilde{N}(d s, d u, d z) \tag{2.46}
\end{equation*}
$$

Then the Doleans-Dade exponential $\mathcal{E}(U)$ is a martingale under $P$ and the probability measure defined by:

$$
\begin{equation*}
\left.\frac{d \mathbb{Q}}{d \mathbb{P}}\right|_{\mathcal{F}_{t}}=\mathcal{E}(U) \tag{2.47}
\end{equation*}
$$

Moreover under $\mathbb{Q}$ the processes $Y_{i}$ are independent on each other and for each $i$ they are CBI processes with parameters ( $a_{i}^{\prime}, b_{i}^{\prime}, \sigma_{i}^{\prime}, \gamma_{i}^{\prime}, \mu_{i}^{\prime}$ ) given by:

$$
\begin{aligned}
& a_{i}^{\prime}=a_{i}-\sigma_{i} \eta_{i}-\gamma_{i} \int_{0}^{\infty} z\left(e^{-\theta_{i} z}-1\right) \mu_{i}(d z) \\
& b_{i}^{\prime}=\frac{a_{i} b_{i}}{a_{i}^{\prime}} \\
& \sigma_{i}^{\prime}=\sigma_{i} \\
& \gamma_{i}^{\prime}=\gamma_{i} \\
& \mu_{i}^{\prime}(d z)=e^{-\theta_{i} z} \mu_{i}(d z)
\end{aligned}
$$

Proof. The process $\left(Y_{1}, \ldots, Y_{n}, U\right)$ is a time homogeneous affine process ([13] Theorem 2.12). The Doleans-Dade exponential $\mathcal{E}(U)$ is a true martingale by checking that the conditions in ([17], Corollary 3.2) are satisfied, so it defines an equivalent probability measure $\mathbb{Q}$. Note that $Z=\mathcal{E}(U)$ is the unique strong solution of

$$
d Z_{t}=Z_{t} d U_{t}
$$

Then for any function $F \in C^{2}\left(\mathbb{R}_{+}^{n}\right)$ the process:

$$
\begin{aligned}
& Z_{t} F\left(Y_{1}(t), \ldots, Y_{n}(t)\right)-\sum_{i=1}^{n} \frac{\sigma_{i}^{2}}{2} \int_{0}^{t} Z_{s} F_{i i}^{\prime \prime}\left(Y_{1}(s), \ldots, Y_{n}(s)\right) Y_{i}(s) d s \\
& -\sum_{i=1}^{n} \int_{0}^{t} Z_{s} F_{i}^{\prime}\left(Y_{1}(s), \ldots, Y_{n}(s)\right)\left\{a_{i} b_{i}-\left[a_{i}-\sigma_{i} \eta_{i}-\gamma_{i} \int_{0}^{\infty} z\left(e^{-\theta_{i} z}-1\right) \mu_{i}(d z)\right] Y_{i}(s)\right\} \\
& -\sum_{i=1}^{n} \int_{0}^{t} Z_{s} Y_{i}(s) d s \int_{0}^{\infty}\left[F \left(Y_{1}(s-), \ldots, Y_{i}(s-)+\gamma_{i} z, \ldots, Y_{n}(s-)-\right.\right. \\
& -F\left(Y_{1}(s-), \ldots, Y_{n}(s-)+F_{i}^{\prime}\left(Y_{1}(s-), \ldots, Y_{n}(s-) Y_{i}(s-) \gamma_{i} z\right] e^{-\theta_{i} z} \mu_{i}(d z)\right.
\end{aligned}
$$

is a local martingale which implies that under $\mathbb{Q}$ the process $Y_{1}, \ldots, Y_{n}$ are independent of each other and for each $i=1, \ldots, n Y_{i}$ is a CBI process with parameters $\left(a_{i}^{\prime}, b_{i}^{\prime}, \sigma_{i}^{\prime}, \gamma_{i}^{\prime}, \mu_{i}^{\prime}\right)$.

Applying now the result proved in Lemma 2.4 we can provide an explicit expression for the instantaneous forward contract in the present modelling framework. By definition of a forward contract we have that:

$$
\begin{equation*}
F(\tau, T)=E^{\mathbb{Q}}\left[S(T) \mid \mathcal{F}_{\tau}\right] \tag{2.48}
\end{equation*}
$$

The model parameters that we used so far were those defined by the hystorical dynamics in such a way that $a_{i}, b_{i}$ and $\mu_{i}$ denote respectively the mean-reversion speed, the long term value and the jump measure with respect to $\mathbb{P}$. Applying the equivalent change of probability measure described in Proposition 2.1 we move to parameters $a_{i}^{\prime}, b_{i}^{\prime}$ and $\mu_{i}^{\prime}$ with respect to $\mathbb{Q}$.
Since the factors $Y_{i}$ remains CBI processes also under $\mathbb{Q}$ we can apply the result given by Lemma 2.4:

Proposition 2.2. The price of the forward contract written on $S$ is given by:

$$
\begin{equation*}
\left.F(\tau, T)=S(\tau)+(\alpha(T)-\alpha(\tau))+\sum_{i=1}^{n}() e^{-a_{i}^{\prime}(T-\tau)}-1\right)\left(Y_{i}(\tau)-b_{i}^{\prime}\right) \tag{2.49}
\end{equation*}
$$

Proof. Applying Lemma 2.4 we have that under the probability measure $\mathbb{Q}$ :

$$
\begin{aligned}
F(\tau, T) & =\alpha(T)+\sum_{i=1}^{n} E^{\mathbb{Q}}\left[Y_{i}(T) \mid \mathcal{F}_{t}\right] \\
& =\alpha(T)+\sum_{i=1}^{n} b_{i}^{\prime}+\left(Y_{i}(\tau)-b_{i}^{\prime}\right) e^{-\alpha_{i}^{\prime}(T-\tau)}
\end{aligned}
$$

and in particular:

$$
S(\tau)=F(\tau, \tau)=\alpha(\tau)+\sum_{i=1}^{n} b_{i}^{\prime}+\left(Y_{i}(\tau)-b_{i}^{\prime}\right)
$$

Hence:

$$
\begin{aligned}
F(\tau, T) & =S(\tau)+F(\tau, T)-F(\tau, \tau) \\
& =\alpha(T)-\alpha(t)+S(\tau)+\sum_{i=1}^{n}\left(e^{-a_{i}^{\prime}(T-\tau)}-1\right)\left(Y_{i}(\tau)-b_{i}^{\prime}\right)
\end{aligned}
$$

### 2.5 The forward model based on Hawkes processes

In this section we will provide an alternative model where our factor are assumed to follow an Hawkes-type dynamic. Let $n$ the number of factors in our model and let $c_{i}, \sigma_{i} \in \mathbb{R}^{+}$
with $i=1, \ldots, n$ constant parameters. We assume that under the historical probability measure $\mathbb{P}$ the dynamics of the $X_{i}$ is given by the following SDE:

$$
\begin{equation*}
X_{i}(t, T)=X_{i}(0, T)-\int_{0}^{t} c_{i} X_{i}(s, T) d s+\sigma_{i} \int_{0}^{t} \sqrt{X_{i}(s, T)} d W_{i}(s)+\int_{0}^{t} \int_{0}^{\infty} z \tilde{J}_{i}(d z, d s) \tag{2.50}
\end{equation*}
$$

where $\tilde{J}_{i}(d z, d s)$ are compensated marked point process with intensity $\lambda_{i}(t)$ satisfying the SDE:

$$
\begin{equation*}
\lambda_{i}(t)=\lambda_{i}(0)-\beta_{i} \int_{0}^{t} \lambda_{i}(s) d s+\alpha_{i} \int_{0}^{t} \int_{0}^{\infty} z J_{i}(d s, d z) \tag{2.51}
\end{equation*}
$$

where in the previous equation $\beta_{i}$ is the rate of exponential decay of the influence of the previous jumps on the intensity level and $\alpha_{i}$ is the amplitude of the memory kernel of each factor $X_{i}$. We also assume that the jump size is distributed according to an exponential density with parameter $\delta_{i}$ for each $\left(\lambda_{i}, X_{i}\right)$ so we can write:

$$
\begin{align*}
\tilde{J}_{i}(d s, d z) & =J_{i}(d s, d z)-\lambda_{i}(s) \mu(d z) d s \\
& =J_{i}(d s, d z)-\lambda_{i}(s) \delta_{i} \exp \left(-\delta_{i} z\right)(d z) d s \tag{2.52}
\end{align*}
$$

We observe, as a remark, that the choice of a square root process for the diffusion part of the forward curves dynamics is motivated by the positivity requirement as well as the choice of the exponential distribution for the sizes of the jumps.
In view of this preliminary work we obtain that formulation for the price of the forward contract written on $n$ underlying factors $X_{i}$ :

$$
\begin{align*}
F(t, T) & =\Lambda(t)-\Lambda(0)+f(0, T)-\sum_{i=1}^{n} \int_{0}^{t} c_{i} X_{i}(s, T) d s+\sum_{i=1}^{n} \int_{0}^{t} \sigma_{i} \sqrt{X_{i}(s, T)} d W_{i}(s) \\
& +\sum_{i=1}^{n} \int_{0}^{t} \int_{0}^{\infty} z \tilde{J}_{i}(d s, d z) \tag{2.53}
\end{align*}
$$

An alternative representation for the for the Hawkes type dynamic given by Dawson and Li , under the Historical probability measure $\mathbb{P}$ is given by the following SDE:

$$
\begin{align*}
X_{i}(t, T) & =X_{i}(0, T)-\int_{0}^{t} c_{i} X_{i}(s, T) d s+\int_{0}^{t} \int_{0}^{X_{i}(s, T)} \sigma_{i} W_{i}(d u, d s) \\
& +\int_{0}^{t} \int_{0}^{X_{i}\left(s_{-}, T\right)} \int_{\mathbb{R}^{+}} z \tilde{N}_{i}(d z, d u, d s) \tag{2.54}
\end{align*}
$$

Also in this case the intensity evolves according to equation (2.51). It appears immediate that the dynamics described by the two models look almost identical when written in the Dawson-Li representation and the unique difference is given by the equation governing the evolution of the intensity process.
At this point in order to have a description with respect to a risk neutral probability measure $\mathbb{Q}$ we need to introduce a measure change. The following proposition provides a measure change which preserves the Hawkes type dynamic.

Proposition 2.3. Let $\left(\lambda_{i}, X_{i}\right)$ be described by equations (2.51) and (2.50) under the historical probability $\mathbb{P}$. Fix $(\eta, \xi) \in \mathbb{R} \times\left(-\delta_{i}, \infty\right)$ and define:

$$
\begin{equation*}
U_{t}:=\sum_{i=1}^{n} \eta_{i} \sigma_{i} \int_{0}^{t} \int_{0}^{X_{i}(s, T)} \sigma_{i} W_{i}(d u, d s)+\sum_{i=1}^{n} \int_{0}^{t} \int_{0}^{X_{i}(s-, T)} \int_{\mathbb{R}^{+}}\left(e^{-\xi_{i} z}-1\right) \tilde{N}_{i}(d z, d u, d s) \tag{2.55}
\end{equation*}
$$

Then the Doleans-Dade exponential $\mathcal{E}(U)$ is a martingale under $\mathbb{P}$ and the probability measure defined by:

$$
\begin{equation*}
\left.\frac{d \mathbb{Q}}{d \mathbb{P}}\right|_{\mathcal{F}_{t}}:=\mathcal{E}(U)_{t} \tag{2.56}
\end{equation*}
$$

is equivalent to $\mathbb{P}$. Moreover the dynamic with respect to $\mathbb{Q}$ takes the following form:

$$
\begin{aligned}
& X_{i}(t, T)=X_{i}(0, T)+\int_{0}^{t} \int_{0}^{X_{i}(s, T)} \sigma_{i}^{\prime} W_{i}(d u, d s)+\int_{0}^{t} \int_{0}^{X_{i}(s-, T)} z \tilde{N}_{i}^{\prime}(d z, d u, d s) \\
& \lambda_{i}(t)=\lambda_{i}(0)-\int_{0}^{t} \beta_{i}^{\prime} \lambda(s) d s+\alpha_{i}^{\prime} \int_{0}^{t} \int_{0}^{\infty} \exp \left(-\beta_{i}^{\prime}(t-s)\right) J_{i}^{\prime}(d s, d z)
\end{aligned}
$$

(the symbol' denotes the parameters under the probability measure $\mathbb{Q}$ ) where:

$$
\begin{aligned}
& c_{i}^{\prime}=c_{i}-\sigma_{i} \eta_{i}-\int_{0}^{\infty} z\left(e^{-\theta_{i} z}-1\right) \mu_{i}(d z) \\
& \sigma_{i}^{\prime}=\sigma_{i} \\
& \alpha_{i}^{\prime}=\alpha_{i} \\
& \beta_{i}^{\prime}=\beta_{i} \\
& \mu_{i}^{\prime}(d z)=e^{-\theta_{i} z} \mu_{i}(d z)
\end{aligned}
$$

We end this section with some remarks:

- parameters $\eta_{i}$ and $\xi_{i}$ can be interpreted as the market price of risk associated with the diffusion/jump part of the $i$-th factor $X_{i}$.
- We assume that the de-seasonalized dynamics of $X_{i}$ is a local martingale under $\mathbb{Q}$ which makes automatically the mean reversion speed $c_{i}=0$ under $\mathbb{Q}$.
- From the formulas in the previous lines, specifying the relations between the model parameters under the risk-neutral measure $\mathbb{Q}$ and the historical measure $\mathbb{P}$, it is clear that in the Hawkes modeling framework a mean reversion speed coefficient $c_{i}$ can be nonzero under $\mathbb{P}$ and zero under $\mathbb{Q}$.


## Chapter 3

## Financial model

In this chapter we discuss an application of the previous two types of processes to power markets; in particular we want to give an estimate of forward curves starting from future prices and then use this forward curves to test the two models based on CBI and Hawkes processes. For this chapter we mainly refer to [8] and also to [2] for the part related on forward curves estimation.

### 3.1 From futures prices to forward curves

In this section we focus our attention on how to determine the historical forward prices starting from future prices observed in the market. First of all we need to set a definition:

Definition 3.1. The price at time $t \geq 0$ of a futures contract with delivery period $\left[T_{1}, T_{2}\right]$ with $t \leq T_{1} \leq T_{2}$ is given by:

$$
F\left(t, T_{1}, T_{2}\right)=\frac{1}{T_{2}-T_{1}} \int_{T_{1}}^{T_{2}} f(t, x) d x
$$

where $f(t, x)$ is the price of the forward contract to be paid upon delivery.
In our modeling framework the value at time $t$ of a future contract with delivery period [ $T_{1}, T_{2}$ ] is given by:

$$
\begin{equation*}
F\left(t, T_{1}, T_{2}\right)=\frac{1}{T_{2}-T_{1}} \int_{T_{1}}^{T_{2}} f(t, x) d x=\Lambda(t)-\Lambda(0)+\frac{1}{T_{2}-T_{1}}\left(\sum_{i=1}^{n} \int_{T_{1}}^{T_{2}} X_{i}(t, x) d x\right) \tag{3.1}
\end{equation*}
$$

Introducing now the dynamics of factors $X_{i}$ in Equation (3.1) under the risk neutral probability $\mathbb{Q}$, we obtain that in the CBI framework:

$$
\begin{align*}
F\left(t, T_{1}, T_{2}\right) & =\Lambda(t)-\Lambda(0)+\frac{1}{T_{2}-T_{1}} \int_{T_{1}}^{T_{2}} f(0, x) d x \\
& +\frac{1}{T_{2}-T_{1}} \sum_{i=1}^{n} \sigma_{i} \int_{T_{1}}^{T_{2}} \int_{0}^{t} \int_{0}^{X_{i}(s, x)} W_{i}(d s, d y) d x  \tag{3.2}\\
& +\frac{1}{T_{2}-T_{1}} \sum_{i=1}^{n} \gamma_{i} \int_{T_{1}}^{T_{2}} \int_{0}^{t} \int_{0}^{X_{i}(s-, x)} \int_{\mathbb{R}^{+}} z \tilde{N}_{i}(d s, d y, d z) d x .
\end{align*}
$$

and in the Hawkes setting:

$$
\begin{align*}
F\left(t, T_{1}, T_{2}\right) & =\Lambda(t)-\Lambda(0)+\frac{1}{T_{2}-T_{1}} \int_{T_{1}}^{T_{2}} f(0, x) d x \\
& +\frac{1}{T_{2}-T_{1}} \sum_{i=1}^{n} \sigma_{i} \int_{T_{1}}^{T_{2}} \int_{0}^{t} \sqrt{X_{i}(s, x)} d W_{i}(d s) d x  \tag{3.3}\\
& +\frac{1}{T_{2}-T_{1}} \sum_{i=1}^{n} \gamma_{i} \int_{T_{1}}^{T_{2}} \int_{0}^{t} \int_{\mathbb{R}^{+}} z \tilde{J}_{i}(d z, d s) d x .
\end{align*}
$$

In order to avoid arbitrage opportunities the prices of futures with different delivery period must satisfy some specific time consistency relations. In particular if we have a contract with delivery period $\left[T_{1}, T_{n}\right]$ it is linked to the values of the contract in intervals $\left[T_{i}, T_{i}+1\right]$ with $i=1, \ldots, n-1$ by the following:

$$
\begin{equation*}
F\left(t, T_{1}, T_{n}\right)=\frac{1}{T_{n}-T_{1}} \sum_{i=1}^{n-1}\left(T_{i+1}-T_{i}\right) F\left(t, T_{i}, T_{i+1}\right) \tag{3.4}
\end{equation*}
$$

From a theoretical point of view the contracts are settled continuously over the delivery period but in practice they are settled at discrete times; assuming settlement at $N$ points in time $u_{1}<u_{2}<\cdots<u_{n}$ with $u_{1}=T_{1}$ and $u_{N}=T_{2}$, thus the discrete version of Equation (3.1) becomes:

$$
\begin{equation*}
F\left(t, T_{1}, T_{2}\right)=\frac{1}{T_{2}-T_{1}} \sum_{i=1}^{N} f\left(t, u_{i}\right) \Delta_{i} \tag{3.5}
\end{equation*}
$$

with $\Delta_{i}=u_{i+1}-u_{i}$. The main goal is to provide a forward dynamics formulation starting from the futures prices that we observe in the market.

Notation 3.1. We use $T_{i}^{s}$ and $T_{i}^{e}$ to denote respectively the first and the last day of the delivery period for the $i$-th contract

From the market we observe the quantity $F\left(0, T^{s}, T^{e}\right)$ for every contract with different possible choices of $T^{s}$ and $T^{e}$ with $T^{e}-T^{s}=7,30,90$ and 365 days according to the type of the contract and where 0 denotes the current date; in other words for any day in an historical horizon we observe the value $F\left(0, T^{s}, T^{e}\right)$ which corresponds to the price of a future contract with delivery period $\left[T^{s}, T^{e}\right]$.
The initial condition that we need to use the Heat-Jarrow-Morton approach when modeling forward is a smooth curve describing today forward prices which must be extracted from futures prices observed in the market. Following the approach given in [2] we set:

Assumption 3.1. The forward curve can be rapresented as the sum of two continuous functions $\Lambda(u)$ and $\varepsilon(u)$ :

$$
\begin{equation*}
f(u)=\Lambda(u)-\Lambda(0)+\varepsilon(u), \quad u \in\left[T^{s}, T^{e}\right] \tag{3.6}
\end{equation*}
$$

where $T^{s}$ is the starting day of the settlement period for the contract with the closest delivery period and $T^{e}$ is the first day of the settlement period for the contract with the farthest delivery period.

We interpret $\Lambda(u)$ as a seasonality function and $\varepsilon(u)$ as an adjustment function that captures the forward curve's deviation from the seasonality.
For the seasonality function, following [2], we assume that:

$$
\begin{equation*}
\Lambda(u)=a \cos \left((u-b) \frac{2 \pi}{365}\right) \tag{3.7}
\end{equation*}
$$

where $a \in \mathbb{R}^{+}$is the minimum of the prices all over the contracts and $b \in \mathbb{R}$ is the normalized distance between the end of the last day of the year from the day when the minimum occurs. Following again the approach from [2] we require for the adjustment function that it is twice continuously differentiable and horizontal at time $T^{e}$, that is:

$$
\varepsilon^{\prime}\left(T^{e}\right)=0
$$

This flatness condition is due to the fact that the long end of the curve may be several years ahead and obviously the market's view on risk becomes less and less sensitive as time goes by.

## Definition 3.2. Let:

$$
\begin{equation*}
C_{0}^{2}\left(\left[T^{s}, T^{e}\right]\right)=\left\{\varepsilon:\left[T^{s}, T^{e}\right] \rightarrow \mathbb{R} \text { s.t. } \varepsilon^{\prime \prime} \in C_{0}\left(T^{s}, T^{e}\right) \text { and } \varepsilon^{\prime}\left(T^{e}\right)=0\right\} . \tag{3.8}
\end{equation*}
$$

Furthermore we consider $\mathcal{C}$ the set of polynomial spline functions of order four which belong to $C_{0}^{2}\left(\left[T^{s}, T^{e}\right]\right)$. Then we define the adjustment function on an interval $\left[T^{s}, T^{e}\right]$ the function that minimizes over $\mathcal{C}$ the integral:

$$
\begin{equation*}
\int_{T^{s}}^{T^{e}}\left[\varepsilon^{\prime \prime}(u)\right]^{2} d u . \tag{3.9}
\end{equation*}
$$

To determine the adjustment function, and therefore the smoothest forward curve a detailed algorithm is presented in [8], other approaches can be found in [2].
At this point, applying choosing different initial day (the 0 in the formulation of $f(0, u)$ ) in a fixed (sufficiently long) temporal horizon, we find a certain amount of different forward curves $f(0, u)$ s.t. $u \in\left[T^{s}, T^{e}\right]$ as in Assumption 3.1.
Given this collection of curves we want to detect if there are times such that positive jumps in the price occur. For a fixed maturity $T$ we define:

$$
V_{t}=f(t, T)
$$

If we plot the forward curves previously determined as functions of time to maturity ( $x$ axes) we can visualize the quantity $V_{t}$ as a vertical section of such graph for a fixed time to maturity $T$. Clearly $t$ ranges over the number of curves that we dispose. The simplest way to detect if a jump occurs is to fix a certain $\Theta \in \mathbb{R}^{+}$, then we say that a jump occurs at time $\hat{t}$ if:

$$
\left|V_{\hat{t}+1}-V_{\hat{t}}\right| \geq \Theta .
$$

An other possible approach is given by an iterative least square algorithm. Set $n$ the number of forward curves available and $\mathcal{N}=\{0,1, \ldots, n-1\}$ then the algorithm reads as follows:

1. Set $\sigma_{1}^{2}$ by:

$$
\sigma_{1}^{2}=\frac{1}{n-2} \sum_{t \in \mathcal{N}} \frac{\left(V_{t+1}-V_{t}\right)^{2}}{V_{t}}
$$

2. Define $\mathcal{M}_{1} \subseteq \mathcal{N}$ the set of elements $t$ such that $\frac{V_{t+1}-V_{t}}{\sqrt{V_{t}}} \geq 3 \sigma_{1}$
3. Set $m_{1}=\left|\mathcal{M}_{1}\right|$ and $i=1$
4. While $m_{i} \neq 0$ repeat the computation:

- set $i=i+1$ and then:

$$
k=\sum_{j=1}^{i-1} m_{j}, \quad K=\bigcup_{j=1}^{i-1} \mathcal{M}_{j}
$$

and

$$
\begin{equation*}
\sigma_{i}^{2}=\frac{1}{n-k-1} \sum_{t \in \mathcal{N} \backslash K} \frac{\left(V_{t+1}-V_{t}\right)^{2}}{V_{t}} \tag{3.10}
\end{equation*}
$$

- Define $\mathcal{M}_{i} \subseteq \mathcal{N} \backslash K$ the set of elements $t$ such that $\frac{V_{t+1}-V_{t}}{\sqrt{V_{t}}} \geq 3 \sigma_{i}$
- set $m_{i}=\left|\mathcal{M}_{i}\right|$.

Looking at the algorithm we can notice that, as $i$ increases, the number $\sigma_{i}^{2}$ decreases and so the number of jumps detected; it is reasonable, looking at our data set, to stop the algorithm after a certain amount of iteration in order to focus our attention on larger jumps; if one waits until when $m_{k}=0$ a lot of very small sized jumps occur.

### 3.2 Testing the model

### 3.2.1 Parameter estimation

Before proceeding with the Kolmogorov-Smirnov test (KS) on the two models we need to estimate the sizes of jumps (whose occurrence have been computed in Section 3.1) and the parameters characterizing the drift and the volatility coefficients.
Since we assumed for both CBI and Hawkes model that the size of jumps is distributed as an exponential random variable we can estimate the size of jumps $\delta$ via its maximum likelihood estimator; let $z_{i}$ the size of the $i$-th jump, and $L$ the number of jump estimated with the algorithm at the end of Section 3.1, then the estimator is:

$$
\begin{equation*}
\hat{\delta}=\frac{L}{\sum_{i=1}^{n} z_{i}} . \tag{3.11}
\end{equation*}
$$

At this point to estimate the parameters appearing in the drift coefficient of our forward dynamics $a_{1}$ in Equation (2.43) and $c_{1}$ in Equation (2.50) (recall that we are assuming to work with a unique factor $X$ ) we first identify them by $\tilde{a}$ and then we recall that:

Remark 3.1. Under the historical probability measure $\mathbb{P}$ both the CBI and Hawkes process in discrete form reads as (assuming no seasonality and no jumps):

$$
X(t+1, T)=X(t, T)-\int_{t}^{t+1} \tilde{a} X(s, T) d s
$$

where $X$ is the unique factor in the forward dynamic.
Thus to estimate $\tilde{a}$ we use, for a fixed maturity $T$ :

$$
\begin{equation*}
\hat{\tilde{a}}=1-\frac{X(t+1, T)}{X(t, T)} \tag{3.12}
\end{equation*}
$$

Finally to estimate the volatility parameter $\sigma$ we take:

$$
\begin{equation*}
\hat{\sigma}=\sigma_{2}, \tag{3.13}
\end{equation*}
$$

where $\sigma_{2}$ is given by Equation (3.10) of the algorithm to detect jumps at the end of Section 3.1.

### 3.2.2 KS test for the models

We want now to perform statistical tests on the models described in Chapter 2 and we want to establish which one of the two model proposed (the one based on CBI processes and the other based on Hawkes processes) is better. To perform such tests we will use KolmogorovSmirnov (KS) test, it is a non-parametric goodness-of-fit test and is used to determine whether two distributions differ, or whether an underlying probability distribution differs from a hypothesized distribution. In our case we will test the null hypothesis:
$H_{0}=$ "the distribution of the data is the same of the one coming from the model"
against the alternative $H_{1}$.
Before using this test we need to estimate the intensity from our data set. We have as input the occurrences of jumps $\tau_{1}<\ldots \tau_{n}=T$ then in case of the CBI model we know the stochastic intensity $\lambda(t)$ is proportional to $X(t, T)$ (see [15]) and the constant of proportional is estimated using the number of jumps and the cumulative de-seasonalized forward prices. In case of the Hawkes dynamic to estimate the stochastic intensity we have to determine the values of parameters $\alpha, \beta$ and $\lambda(0)$ in Equation (2.51). The log-likelihood of an Hawkes process is given by:

$$
\begin{equation*}
\log L\left(\tau_{1}, \ldots, \tau_{N}\right)=-\lambda_{0} \tau_{N}+\sum_{i=1}^{N} \frac{\alpha}{\beta}\left(e^{-\beta\left(\tau_{n}-\tau_{i}\right)}-1\right)+\sum_{i=1}^{N} \log (\lambda(0)+\alpha A(i)) \tag{3.14}
\end{equation*}
$$

with $A(i)=\sum_{\tau_{j}<\tau_{i}} e^{-\beta\left(\tau_{i}-\tau_{j}\right)}$ and $A(1)=0$. To determine the values of parameters $\alpha, \beta$ and $\lambda(0)$ we have to compute the maximum of the previous function as a real function of three variables.
At this point everything is set to perform the KS test to see which of the distribution best models the jumps observed in our set of data. This procedure is entirely described in [8]; in particular starting from an historical set of data on French power futures prices and after the implementation of the algorithms to compute forward prices and to estimates jumps on such prices and their sizes, the KS statistical test is performed (with a significance level of 0.05 ) to test the to models based on CBI and Hawkes processes and an additional toy model based on Poisson processes.


Figure 3.1: This figure shows the evolution of gold future prices (red line) and estimated jump times (blue vertical lines)

### 3.3 Focus on jump detection

In this subsection we focus our attention on the algorithm to estimate the number of jumps presented ad the end of Section 3.1 and we apply it to three different set of data:

- A simulated geometric Brownian motion trajectory with jumps given by an Hawkes process
- an historical set of data on gold futures ${ }^{1}$
- an historical set of data on crude oil futures ${ }^{2}$

In particular starting from this set of data (the first one is a toy model) we want first to apply the algorithm (whose code is available in Appendix C) to detect jumps and then we want to see if we can say something about the correlation of this jumps or not.

Figure 3.1 shows the evolution of the future price of gold from the starting date $2017 / 6 / 29$ to $2023 / 4 / 6$ in red line and the jumps in the future prices detected using the algorithm of jump detection (blue line). In particular looking at the position of jumps we see an high concentration of jumps around the first months of 2020, this is due to the beginning of corona-virus pandemic and in particular by the global collapse of the economies as a result of policies of home confinements and the clear contagion effect between markets of very different natures and geographies. After this gold experiment an

[^1]

Figure 3.2: This figure shows the evolution of crude oil (red line) and estimated jump times (blue vertical lines)


Figure 3.3: This figure shows the evolution of a simulated geometric Brownian motion with Hawkes jumps and the jumps detected with the algorithm
exponential recovery and the major driver behind the strong purchasing of Gold during the year 2020 lies in its store of value nature. As time goes by just few jumps are experimented but without an apparently fixed pattern.
Figure 3.2 shows the evolution of the future prices of crude oil from the starting date 2017/11/21 to 2023/4/6 (red line) and the jumps in the price detected using the algorithm of jump detection (blue line). We can see that around February/March 2020 there is a big downward jump due to corona-virus pandemic and the consequent global collapse of the economies while an high concentration of jumps, due to the explosion of the war between Ukraine and Russia, is present starting from the first months of 2022.

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## Appendix A

## Useful results

## A. 1 Doubly Stochastic Or Conditional Poisson Processes

A nonmathematical definition of a doubly stochastic Poisson process consists in describing a two step randomization procedure: in the first step one draws at random the trajectory of a driving process say $Y_{t}$ and once the trajectory is selected one generates a Poisson process of intensity $f\left(t, Y_{t}\right)$. Of course such an intensity is random, since $Y_{t}$ is a random process and thereforethe point process is Poisson only conditionally with respect to $Y_{t}$.

Definition A.1. Let $\left\{N_{t}\right\}_{t \geq 0}$ a point process adapted to a $\sigma$-algebra $\left\{\mathcal{F}_{t}\right\}_{t \geq 0}$ and let $\left\{\lambda_{t}\right\}_{t \geq 0}$ a nonnegative measurable process. Suppose that:

$$
\lambda_{t} \text { is } \mathcal{F}_{0}-\text { measurable } \forall t \geq 0
$$

and that:

$$
\int_{0}^{t} \lambda_{s} d s<\infty P-a . s . \quad \forall t \geq 0
$$

If for all $0 \leq s \leq t$ and $u \in \mathbb{R}$ :

$$
\begin{equation*}
E\left[e^{i u\left(N_{t}-N_{s}\right)} \mid \mathcal{F}_{s}\right]=\exp \left\{\left(e^{i u}-1\right) \int_{s}^{t} \lambda_{v} d v\right\} \tag{A.1}
\end{equation*}
$$

then $N_{t}$ is called a $\left(P, \mathcal{F}_{t}\right)$-doubly stochastic Poisson process.
In particular if $\lambda_{t}$ is deterministic then $N_{t}$ is a $\left(P, \mathcal{F}_{t}\right)$-Poisson process. As an example we consider the case in which $\lambda_{t}=f\left(t, Y_{t}\right)$ for some properly measurable and non-negative function $f$ and for some measurable process $Y_{t}$ and if $\mathcal{F}_{0}$ contains $\mathcal{F}_{\infty}^{Y}$ then $N_{t}$ is a doubly stochastic Poisson process driven by $Y_{t}$. We mention now an important theorem which gives a characterization for doubly stochastic Poisson processes:

Theorem A.1. Let $\left\{N_{t}\right\}_{t \geq 0}$ a point process adapted to some history $\left\{\mathcal{F}_{t}\right\}_{t \geq 0}$ and $\left\{\lambda_{t}\right\}_{t \geq 0}$ a non negative process such that $\forall t \geq 0$ :

1. $\lambda_{t}$ is $\mathcal{F}_{t}$-measurable.
2. $\int_{0}^{t} \lambda_{s} d s<\infty P$-a.s.

Then if the equality:

$$
\begin{equation*}
E\left[\int_{0}^{\infty} C_{s} d N_{s}\right]=E\left[\int_{0}^{\infty} C_{s} \lambda_{s} d s\right] \tag{A.2}
\end{equation*}
$$

Is verified for all nonnegative $\mathcal{F}_{t}$ predictable processes $C_{t}$ then $N_{t}$ is a doubly stochastic Poisson process with the $\mathcal{F}_{t}$-intensity $\lambda_{t}$.

## A. 2 Multivariate and Marked Point Processes

Let $\left\{T_{n}\right\}_{n \in \mathbb{N}}$ a univariate point process and $\left\{Y_{n}\right\}_{n \in \mathbb{N}}$ a sequence of random variables with values in $\{1,2, \ldots, K\}$ all defined in $(\Omega, \mathcal{F}, P)$. For each $k$ we can consider:

$$
N_{t}(k)=\sum_{n \in \mathbb{N}} \mathbf{1}_{\left\{T_{n} \leq t\right\}} \mathbf{1}_{\left\{Y_{n}=k\right\}}
$$

That is a random variable which counts the number of jumps of length $k$ in the interval $(0, t]$. Moreover each $N_{t}(k)$ is a Poisson process and $\forall k \neq h$ the processes $N_{t}(k)$ and $N_{t}(h)$ have no jumps in common. We have two possible representation to describe this process called multivariate Poisson process; either with $\left(T_{n}, Y_{n}\right)_{n \in \mathbb{N}}$ or with a $K$-dimensional vector $\left(N_{t}(1), \ldots, N_{t}(K)\right)_{t \geq 0}$. Considering the first representation $\left(T_{n}, Y_{n}\right)_{n \in \mathbb{N}}$ we may interpret $T_{n}$ as the time of occurrence of some event and $Y_{n}$ as an attribute or a mark of that event.

Definition A.2. An E-marked point process is a double sequence $\left(T_{n}, Y_{n}\right)_{n \in \mathbb{N}}$ where:

1. $T_{n}$ is a univariate point process.
2. $Y_{n}$ is a sequence of random variables with values on $(E, \mathcal{E})$.

Generalizing the previous representation of a multivariate point process we associate for any measurable set $A \in \mathcal{E}$ the counting process:

$$
N_{t}(A)=\sum_{n \in \mathbb{N}} \mathbf{1}_{\left\{T_{n} \leq t\right\}} \mathbf{1}_{\left\{Y_{n} \in A\right\}}
$$

And we let simply $N_{t}:=N_{t}(E)$. Then we consider the filtration

$$
\mathcal{F}_{t}^{N}=\sigma\left\{N_{s}(A): s \leq t, A \in \mathcal{E}\right\}
$$

And we define the counting measure:

$$
p((0, t], A)=N_{t}(A) t \geq 0, A \in \mathcal{E}
$$

this measure is $\sigma$-finite under the assumption of non explosivity of $T_{n}$. Moreover it allows to obtain more coincise expressions for integrals of the type:

$$
\begin{equation*}
\int_{0}^{t} \int_{E} H(s, y) p(d s, d y)=\sum_{n \in \mathbb{N}} H\left(T_{n}, Y_{n}\right) \mathbf{1}_{\left\{T_{n} \leq t\right\}} \tag{A.3}
\end{equation*}
$$

Assume now that for each $A \in \mathcal{E}$ we have that the point process $N_{t}(A)$ admits the intensity $\lambda_{t}(A)$; this leads to a measure valued intensity $\lambda_{t}(d y)$, we can then generalize equation (A.2):

$$
\begin{equation*}
E\left[\int_{0}^{\infty} \int_{E} H(s, y) p(d s, d y)\right]=E\left[\int_{0}^{\infty} \int_{E} H(s, y) \lambda_{s}(d y) d s\right] \tag{A.4}
\end{equation*}
$$

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That has to be valid for all nonnegative $\mathcal{F}_{t}$ predictable $E$-marked process where in this case $\mathcal{F}_{t}$-predictable means measurable w.r.t. $\mathcal{P}\left(\mathcal{F}_{t}\right) \otimes \mathcal{E}$ where $\mathcal{P}\left(\mathcal{F}_{t}\right)$ is the predictable $\sigma$-algebra on $(0, \infty) \times \Omega$. We have also the generalization of (1.37) in the form:

$$
\begin{equation*}
q(d s, d y)=p(d s, d y)-\lambda_{s}(d y) d s \tag{A.5}
\end{equation*}
$$

where $q(d s, d y)$ is a signed measure valued martingale in the sense that:

$$
\int_{0}^{t} \int_{E} H(s, y) q(d s, d y)
$$

is a $\left(P, \mathcal{F}_{t}\right)$ martingale for each $\mathcal{F}_{t}$ predictable $E$-marked process satisfying appropriate integrability conditions. The most common form of intensity is:

$$
\lambda_{t}(d y)=\lambda_{t} m_{t}(d y)
$$

where $\lambda_{t}$ is non negative and $\mathcal{F}_{t}$-predictable and represents the intensity of the process $N_{t}(E)$ and $m_{t}(d y)$ is a probability measure on $E$. The pair $\left(\lambda_{t}, m_{t}(d y)\right)$ is called the $\left(P, \mathcal{F}_{t}\right)$-characteristic.

## A. 3 Martingale representation and generalized Ito formula

Martingale representation results are widely used in finance especially when it comes to solving hedging problems. For pure "Wiener-martingales" we have in fact the well known result that every square integrable martingale with respect to a filtration generated by a Wiwener process is, up to an additive constant, a stochastic integral of the Ito type. We recall now an important theorem:

Theorem A.2. Let $\left(\Omega, \mathcal{F}, \mathcal{F}_{t}, P\right)$ be a probability space where $\mathcal{F}_{t}=\mathcal{F}_{0} \vee \mathcal{F}_{t}^{p}$ with $\mathcal{F}_{t}^{p}$ filtration generated by a marked point process represented by the counting measure $p(d t, d y)$. Then any $\left(P, \mathcal{F}_{t}\right)$-martingale $M_{t}$ admits the representation:

$$
\begin{equation*}
M_{t}=M_{0}+\int_{0}^{t} \int_{E} H(s, y) q(d s, d y) \tag{A.6}
\end{equation*}
$$

with $q(\cdot, \cdot)$ as in (A.5) and $H$ an integrable (with respecto to $\lambda_{t}(d y)$ ) $\mathcal{F}_{t}$-predictable $E$ marked process. This representation is essentially unique.

Now using the definition of a marked point process and of integrals in the form of (A.3) we may now consider processes of the general type:

$$
\begin{equation*}
X_{t}=X_{0}+\int_{0}^{t} \alpha_{s} d s+\int_{0}^{t} \beta_{s} d w_{s}+\int_{0}^{t} \int_{E} \gamma(s, y) p(d s, d y) \tag{A.7}
\end{equation*}
$$

that are called jump diffusion processes and where the coefficients satisfy the implicit integrability condition, $\beta_{t}$ is $\mathcal{F}_{t}$ adapted and $\gamma_{t}$ is predictable in the sense of the previous definition. We may rewrite the previous equation in differential form as:

$$
\begin{equation*}
d X_{t}=X_{t-}\left(\alpha_{t} d t+\beta_{t} d w_{t}+\int_{E} \gamma(t, y) p(d t, d y)\right) \tag{A.8}
\end{equation*}
$$

In the previous equation $t$ - specifies the predictability requirement on the last term. As we saw before the last term can be rewritten as:

$$
\begin{equation*}
\int_{E} \gamma(t, y) p(d t, d y)=\gamma\left(t, Y_{t}\right) d N_{t} \tag{A.9}
\end{equation*}
$$

where $N_{t}=N_{t}(E)=p((0, t], E)$ is the total number of jumps and $Y_{t}$ denotes the piecewise constant, left continuous time interpolation of the sequence $Y_{n}$. A particular solution to (A.7) is given by the following exponential formula:

$$
\begin{align*}
X_{t} & =X_{0} \exp \left\{\int_{0}^{t}\left(\alpha_{s}-\frac{1}{2} \beta_{s}^{2}\right) d s+\int_{0}^{t} \beta_{s} d w_{s}+\int_{0}^{t} \log \left(1+\gamma\left(s, Y_{s}\right)\right) d N_{s}\right\} \\
& =X_{0} \exp \left\{\int_{0}^{t}\left(\alpha_{s}-\frac{1}{2} \beta_{s}^{2}\right) d s+\int_{0}^{t} \beta_{s} d w_{s}\right\} \prod_{n=1}^{N_{t}}\left(1+\gamma\left(T_{n}, Y_{n}\right)\right) \tag{A.10}
\end{align*}
$$

While the diffusion part in the expression follows from the usual Ito formula the jump part follows from the so called exponential formula of Lebesgue-Stieltjes Calculus but can be also obtained from the generalized Ito formula.
Let $X_{t}$ a process satisfying (A.7) and let $F \in C^{1,2}$ function $F\left(t, X_{t}\right)$; the generalized Ito formula:

$$
\begin{align*}
d F\left(t, X_{t}\right)= & F_{t}(\cdot) d t+F_{X}(\cdot) \alpha_{t} d t+\frac{1}{2} F_{X X}(\cdot) \beta_{t}^{2} d t+F_{X}(\cdot) \beta_{t} d w_{t}  \tag{A.11}\\
& +\left[F\left(t, X_{t-}+\gamma\left(t, Y_{t}\right)\right)-F\left(t, X_{t-}\right)\right] d N_{t}
\end{align*}
$$

that in the specific case of (A.8) becomes:

$$
\begin{align*}
d F\left(t, X_{t}\right)= & F_{t}(\cdot) d t+F_{X}(\cdot) X_{t} \alpha_{t} d t+\frac{1}{2} F_{X X}(\cdot) X_{t}^{2} \beta_{t}^{2} d t+F_{X}(\cdot) X_{t} \beta_{t} d w_{t}  \tag{A.12}\\
& +\left[F\left(t, X_{t-}\left(1+\gamma\left(t, Y_{t}\right)\right)\right)-F\left(t, X_{t-}\right)\right] d N_{t}
\end{align*}
$$

and where again $N_{t}=N_{t}(E)=p((0, t], E)$ and $(\cdot)$ stands for $\left(t, X_{t}\right)$ the pedices denotes partial derivatives. If we write the last equation in the integral form we have two equivalent representations for the last term:

$$
\begin{equation*}
\int_{0}^{t}\left[F\left(s, X_{s-}\left(1+\gamma\left(s, Y_{s}\right)\right)\right)-F\left(s, X_{s-}\right)\right] d N_{s}=\sum_{n=1}^{N_{t}}\left[F\left(T_{n}, X_{T_{n}}\right)-F\left(T_{n}, X_{T_{n}^{-}}\right)\right] \tag{A.13}
\end{equation*}
$$

Now choosing as $F(t, X)=\log X$ we have:

$$
\begin{equation*}
d F=\alpha_{t} d t-\frac{1}{2} \beta_{t}^{2} d t+\beta_{t} d w_{t}+\log \left(1+\gamma\left(t, Y_{t}\right)\right) d N_{t} \tag{A.14}
\end{equation*}
$$

from which:

$$
\begin{equation*}
\log X_{t}=\log X_{0}+\int_{0}^{t}\left(\alpha_{s}-\frac{1}{2} \beta_{s}^{2}\right) d s+\int_{0}^{t} \beta_{s} d w_{s}+\int_{0}^{t} \log \left(1+\gamma\left(s, Y_{s-}\right)\right) d N_{s} \tag{A.15}
\end{equation*}
$$

And taking exponential on both sides we reobtain (A.10).

## Appendix B

## Hawkes process simulation

In this section we discuss the algorithm to simulate an Hawkes process and we provide also the matlab code of such simulation. The main references are [13] and [24].
This algorithm is able to exactly generate the point process and the intensity process by sampling inter-arrival times directly via the underlying analytic distribution functions without numerical inverse and hence avoids simulating intensity paths and introducing discretization bias.

## B. 1 The algorithm

Assume to have a one dimensional Hawkes process $\{(N(t), \lambda(t)), t \geq 0\}$ where $\lambda(t)=$ $\lambda\left(t \mid \mathcal{H}_{t}\right)$ conditional on $\lambda_{0}$ and $N(0)=0$ with jump size distribution $Y \sim G$ and $K$ jump times $\left\{T_{1}, \ldots, T_{K}\right\}$, the algorithm is based on the following steps:

1. Set the initial conditions $T_{0}=0, \lambda\left(T_{0}\right)=\lambda_{0}>a, N(0)=0$ and $k \in\{0,1,2, \ldots, K-$ $1\}$.
2. Simulate the $(k+1)^{\text {th }}$ inter-arrival time $S_{k+1}$ by:

$$
S_{k+1}= \begin{cases}S_{k+1}^{1} \wedge S_{k+1}^{2} & D_{k+1}>0 \\ S_{k+1}^{2} & D_{k+1}<0\end{cases}
$$

where:

$$
D_{k+1}=1+\frac{\delta \ln U_{1}}{\lambda\left(T_{k}^{+}\right)-a} \quad U_{1} \sim \operatorname{Unif}[0,1]
$$

and:

$$
S_{k+1}^{1}=-\frac{1}{\delta} \ln D_{k+1} \quad S_{k+1}^{2}=-\frac{1}{a} \ln U_{2} \quad U_{2} \sim \operatorname{Unif}[0,1]
$$

3. record the $(k+1)^{\text {th }}$ jump-time $T_{k+1}$ by:

$$
T_{k+1}=T_{k}+S_{k+1}
$$

4. record the change at the jump time $T_{k+1}$ in the process $\lambda(t)$ by

$$
\begin{equation*}
\lambda\left(T_{k+1}\right)=\lambda\left(T_{k+1}^{-}\right)+Y_{k+1} \quad Y_{k+1} \sim G \tag{B.1}
\end{equation*}
$$

where:

$$
\lambda\left(T_{k+1}^{-}\right)=\left(\lambda\left(T_{k}\right)-a\right) e^{-\delta\left(T_{k+1}-T_{k}\right)}+a
$$

5. Record the change at the jump-time $T_{k+1}$ in the point process $N(t)$ by

$$
\begin{equation*}
N\left(T_{k+1}\right)=N\left(T_{k}^{-}\right)+1 \tag{B.2}
\end{equation*}
$$

We prove now that this algorithm produces an Hawkes process with exponential decaying intensity

Proof. Given the $k$-th jump time $T_{k}$ the point process has the intensity $\left\{\lambda(t): T_{k} \leq t<\right.$ $\left.T_{k+1}\right\}$ following the ordinary differential equation (between the times of the $k$-th and the $k+1$-th jumps):

$$
\left\{\begin{array}{l}
\frac{d \lambda(t)}{d t}=-\delta(\lambda(t)-a)  \tag{B.3}\\
\left.\lambda(t)\right|_{t=T_{k}}=\lambda\left(T_{k}\right)
\end{array}\right.
$$

The solution of the above ODE is given by:

$$
\begin{equation*}
\lambda(t)=a+C_{2} e^{-\delta t} \tag{B.4}
\end{equation*}
$$

and imposing theboundary conditions we obtain:

$$
\begin{equation*}
C_{2}=\left(\lambda\left(T_{k}\right)-a\right) e^{\delta T_{k}} \tag{B.5}
\end{equation*}
$$

Finally we obtain that for $T_{k} \leq t \leq T_{k}+S_{k+1}$

$$
\begin{equation*}
\lambda(t)=a+\left(\lambda\left(T_{k}\right)-a\right) e^{-\delta\left(t-T_{k}\right)} \tag{B.6}
\end{equation*}
$$

We want now to evaluate the cumulative distribution function of the $k+1$-th inter arrival time $S_{k+1}$

$$
\begin{aligned}
F_{S_{k+1}}(s) & =P\left(S_{k+1} \leq s\right) \\
& =1-P\left(S_{k+1}>s\right) \\
& =1-P\left(N\left(T_{k}+s\right)-N\left(T_{k}\right)=0\right.
\end{aligned}
$$

Since we know that $\left\{N\left(T_{k}+s\right)-N\left(T_{k}\right), s \geq 0\right\}$ is a Poisson random variable with mean:

$$
E\left(N\left(T_{k}+s\right)-N\left(T_{k}\right)\right)=\int_{T_{k}}^{T_{k}+s} \lambda(u) d u
$$

Therefore we obtain that:

$$
F_{S_{k+1}}(s)=1-e^{-\int_{T_{k}}^{T_{k}+s} \lambda(u) d u}
$$

we make now the change of variable $v=u-T_{k}$ to find:

$$
F_{S_{k+1}}(s)=1-e^{-\int_{0}^{s} \lambda\left(T_{k}+v\right) d v}
$$

and finally replacing the expression obtained for $\lambda$ in equation (B.6) we derive:

$$
\begin{aligned}
F_{S_{k+1}}(s) & =1-e^{-\int_{0}^{s}\left[a+\left(\lambda\left(T_{k}\right)-a\right) e^{-\delta\left(T_{k}+v-T_{k}\right)}\right] d v} \\
& =1-e^{-\int_{0}^{s}\left[a+\left(\lambda\left(T_{k}\right)-a\right) e^{-\delta v}\right] d v} \\
& =1-e^{-\left(\lambda\left(T_{k}\right)-a\right) \frac{1-e^{-\delta s}}{\delta}-a s}
\end{aligned}
$$

At this point we decompose $S_{k+1}$ into two simpler random variables $S_{k+1}^{1}$ and $S_{k+1}^{2}$ via:

$$
\begin{aligned}
& P\left(S_{k+1}^{1}>s\right)=e^{-\left(\lambda\left(T_{k}\right)-a\right) \frac{1-e^{-\delta s}}{\delta}} \\
& P\left(S_{k+1}^{2}>s\right)=e^{-a s}
\end{aligned}
$$

We need now to simulate $S_{k+1}^{1}$ and $S_{k+1}^{2}$ since we have that:

$$
\begin{equation*}
P\left(S_{k+1}>s\right)=P\left(S_{k+1}^{1} \wedge S_{k+1}^{2}>s\right) \tag{B.7}
\end{equation*}
$$

Recall that:

$$
F_{S_{k+1}^{1}}(s)=P\left(S_{k+1}^{1}<s\right)=1-e^{-\left(\lambda\left(T_{k}\right)-a\right) \frac{1-e^{-\delta s}}{\delta}}
$$

then we set:

$$
e^{-\left(\lambda\left(T_{k}\right)-a\right) \frac{1-e^{-\delta S_{k+1}^{1}}}{\delta}}=U_{1}
$$

and we invert explicitely the function:

$$
\begin{aligned}
& e^{-\left(\lambda\left(T_{k}\right)-a\right) \frac{1-e^{-\delta S_{k+1}^{1}}}{\delta}}=U_{1} \\
& \Rightarrow-\left(\lambda\left(T_{k}\right)-a\right) \frac{1-e^{-\delta S_{k+1}^{1}}}{\delta}=\ln U_{1} \\
& \Rightarrow-\frac{1-e^{-\delta S_{k+1}^{1}}}{\delta}=\frac{\ln U_{1}}{\lambda\left(T_{k}\right)-a} \\
& \Rightarrow e^{-\delta S_{k+1}^{1}}=\frac{\ln U_{1}}{\lambda\left(T_{k}\right)-a} \\
& \Rightarrow-\delta S_{k+1}^{1}=1+\frac{\delta \ln U_{1}}{\lambda\left(T_{k}\right)-a} \\
& S_{k+1}^{1}=-\frac{1}{\delta} \ln \left(1+\frac{\delta \ln U_{1}}{\lambda\left(T_{k}\right)-a}\right)
\end{aligned}
$$

Note that the random variable $S_{k+1}^{1}$ may take with positive probability the value $\infty$. Moreover the condition for simulate a valid $S_{k+1}^{1}$ is that:

$$
1+\frac{\delta \ln U_{1}}{\lambda\left(T_{k}\right)-a}>0
$$

thus we introduce the random variable $D_{k+1}$ setting:

$$
D_{k+1}=1+\frac{\delta \ln U_{1}}{\lambda\left(T_{k}\right)-a}
$$

Now for the simulation of $S_{k+1}^{2}$, since we have that $S_{k+1}^{2} \sim \operatorname{Exp}(a)$ we use the standard simulation method:

$$
S_{k+1}^{2}=\frac{1}{a} \ln U_{2}
$$

Hence for the simulation of $S_{k+1}$ we have:

$$
S_{k+1}= \begin{cases}S_{k+1}^{1} \wedge S_{k+1}^{2} & \text { if } D_{k+1}>0 \\ S_{k+1}^{2} & \text { if } D_{k+1}<0\end{cases}
$$

Therefore the $(k+1)$-th jump in the Hawkes process is given by:

$$
T_{k+1}=T_{k}+S_{k+1}
$$

and the change in $\lambda(t)$ and $N(t)$ at time $T_{k+1}$ can be easily derived as given by (B.1) and (B.2).

## B. 2 Code of the simulation

In the sequel the code of the simulation of the algorithm previously described

```
clear all; clc;
% setting initial conditions
tsimul=0;
tmax=100;
npts=500;
a=0.5;
alpha=1;
b=1.3;
k=1;
t=linspace(tsimul,tmax, npts);
Rt=zeros(1, length(t));
Rt(1)=a;
jump=Rt(1);
[Rt,Nt]=hawkesSimulation(tsimul, tmax,a, alpha,b,t);
figure(1)
hold on
title('Intensity Hawkes process')
plot(t,Rt,'r-')
stairs(t,0.05*Nt,'b-')
hold off
figure(2)
hold on
title('Counting process')
stairs(t,Nt,'b-')
hold off
```

This function implements the algorithm and produces two vectors, the first one containing the evolution of the intensity of the process and the second one containing the evolution of the counting process.

```
function [Rt,Nt]=hawkesSimulation(tsimul,tmax,a, alpha,b,t)
Rt=zeros(1,length(t));
```

```
Nt=zeros(1,length(t));
Rt(1)=a;
jump=Rt(1);
k=1;
while tsimul(k,:)<tmax
tjump=tsimul(k,:);
D=1+b*log(rand)/(jump(k,:)-a);
interev=[-log(D)/b,-log(rand)/a];
if D>0
interevNextJump=min(interev);
else
interevNextJump=interev (2);
end
tnextJump=tjump+interevNextJump;
tsimul=[tsimul;tnextJump];
ind=find((tsimul(k,:)<t)&(t<tsimul(k+1,:)));
tind=t(ind);
indbelow=find(t<=tsimul(k,:));
tbelow=t(indbelow);
indabove=find((tsimul(k+1,:)<=t));
tabove=t(indabove);
RtInd=a+(jump(k,:)-a)*exp(-b*(tind-tjump));
RtBelow=zeros(1,length(tbelow));
RtAbove=zeros(1,length(tabove));
Rt=Rt+[RtBelow, RtInd, RtAbove];
Rtmin=a+(jump(k,:)-a)*exp(-b*(tnextJump -tjump));
jump=[jump;Rtmin+alpha];
k=k+1;
end
for i=1:length(tsimul)
iabove=find(t>=tsimul(i,:));
nabove=t(iabove);
ibelow=find(t<tsimul(i,:));
nbelow=t(ibelow);
zer=zeros(1,length(tbelow));
Nt=Nt+[zeros(1,length(nbelow)) ones(1,length(nabove))];
end
```


## Appendix C

## Matlab simulations

In this section we include some Matlab codes where we implement the algorithm to detect jumps described in Section 3.1; in particular we use it to detect jumps in case of a simulated geometric Brownian motion with jumps and in case in which we dispose of historical series of Gold crude oil future prices (data are chosen in the range from 2017/6/29 until 2023/4/6 for gold and from 2017/11/21 until 2023/4/6 for crude oil).

## C. 1 Matlab codes

The following are the main code of the simulation of a geometric Brownian motion with jumps (chosen randomly in the interval $[-5,5]$ ) and the function that performs such computation:

```
% simulation geometric brownian motion with jumps
Tin=0;
Tf=10;
npts=1000;
mu=0.03;
sigma=0.1;
inV=50;
nsimul=20;
lambda=1.5;
t=linspace(Tin,Tf,npts);
a=0.5;
alpha=1;
b=1.3;
k=1;
[Rt,Nt,T]=hawkesSimulation(Tin,Tf,a, alpha,b,t);
[GBM,GBM1,time]=simulateGBMHawkes(inV,Tin,Tf,mu,sigma,npts,T);
hold on
title('Comparison GBM and GBM with Hawkes type jumps')
```

```
plot(time,GBM,'-b')
plot(time,GBM1,'-r')
xline(T(1: end-1))
legend('geometric brownian motion with jumps', ...
'geometric brownian motion', 'time of occurrence of a jump')
hold off
```

```
function [S,S1,T]=simulateGBMHawkes(inV,Tin,Tf,mu,sigma,npts,times)
t=sort(Tin+rand(1,npts)*(Tf-Tin));
T=t;
%t=linspace(Tin,Tf,npts);
r=normrnd (0,1,[1,length(t)]);
S=zeros(1,length(t));
S1=zeros(1, length(t));
S(1)=inV;
S1(1)=inV;
j=1;
for i=2:length(t)
S(i)=S(i-1)*exp ((mu-sigma^2/2)*(t(i)-t(i-1))+...
sigma*sqrt(t(i)-t(i-1))*r(i));
S1(i)=S1(i-1)*exp((mu-sigma^2/2)*(t(i)-t(i-1))+\ldots
sigma*sqrt(t(i)-t(i-1))*r(i));
if (times(j)<t(i))&& (times(j)>=t(i-1))
jump=-5+rand*10;
S(i)=S(i)+jump;
j=j+1;
end
end
```

The following codes are the main code and the function to detect jumps on the set of data that we dispose (simulated GBM with jumps, gold future prices and crude oil future prices).

```
clear all; clc;
%JUMP DETECTION ALGORITHM
%Construction of the futures price vector (random entries)
% n=4000;
% xmin=90;
% xmax=95;
% x =xmin +rand (1,n)*(xmax-xmin);
```

```
% POSSIBLE CHOOSING AS COMMODITY THE SPOT PRICE
% OF GOLD AND CRUDE OIL
% gold=futureprices("futures_gold1.xlsx", "Foglio4",...
% [2, Inf]);
% comm=fliplr(gold')';
% date=readtable("date_gold.xlsx");
% oil=futureprices("futures_gold1.xlsx", "Foglio5", ...
% [2, Inf]);
% comm=fliplr(oil')';
% date=readtable("date_imp.xlsx");
dat=table2array(date);
% SPOT PRICE GENERATED SIMULATING A GBM WITH
% POISSONIAN JUMPS
% setting data to simulate a GBM with poissonian jumps
% Tin=0;
% Tf=10;
%npts=1000;
% mu=0.03;
% sigma=0.1;
% inV=50;
% lambda=1.5;
% [interev,times]=homoPoisson(Tf,lambda);
% [comm,t]= ...
% simulateGBMPoisson(inV,Tin,Tf,mu,sigma,npts,times);
% SPOT PRICE GENERATED SIMULATING A GBM WITH HAWKES JUMPS
% Tin=0;
% Tf=10;
% npts=1000;
% mu=0.03;
% sigma=0.1;
% inV=50;
% nsimul=20;
% lambda=1.5;
% t=linspace(Tin,Tf,npts);
% a=0.5;
% alpha=1;
% b=1.5;
% k=1;
% [Rt,Nt,T]=hawkesSimulation(Tin,Tf,a, alpha,b,t);
% [comm, comm1,time]=...
% simulateGBMHawkes(inV,Tin,Tf,mu,sigma,npts,T);
```

[jumptimes, numberjumpstep, sigmavec]=jumpDetect(comm);

```
hold on
plot(dat,comm,'-r')
xline(dat(jumptimes(1: numberjumpstep (1))),'--b')
hold off
% print -depsc CrudeDil.eps
% print -depsc Gold.eps
% print -depsc GBM.eps
```

function [jumptimes, numberjumpstep, sigmavec]=jumpDetect (comm)
par=3;
$\mathrm{x}=\mathrm{comm}$;
$y=1: 1: \operatorname{length}(c o m m)$;
MatComm=[x; y];
$\mathrm{n}=\mathrm{length}(\mathrm{x})$;
sigma2=0;
M1 = [];
jumptimes=[];
numberjumpstep=[];
sigmavec=[];
Vt=zeros (2, length (comm)-1)';
for $i=1: n-1$
sigma $2=\operatorname{sigma} 2+1 /(n-2) * \ldots$
(MatComm (1, i+1)-MatComm(1,i))~2/MatComm(1,i);
end
for $j=1: n-1$
test=abs((MatComm(1,j+1)-MatComm(1,j))/sqrt(MatComm(1,j)));
$\operatorname{Vt}(j, 1)=j$;
Vt (j, 2) =test;
if test>=par*sqrt(sigma2)
M1 $=[$ M1, $\operatorname{MatComm}(2, j)] ;$
end
end
jumptimes = [jumptimes, M1];
m1=length (M1) ;
numberjumpstep=[numberjumpstep, m1];
sigmavec=[sigmavec,sqrt(sigma2)];

```
% updating x vector
MatComm(:, M1)=[];
k=1;
% counts the number of jumps at each step
counter=0;
while m1 ~= 0
sigma2=0;
M1 = [];
M2 = [];
k=k+1;
counter=counter +m1;
for i=1:length(MatComm)-1
sigma2=sigma2+1/(n-counter - 1) *(MatComm(1,i+1) ...
-MatComm(1,i))^2/MatComm(1,i);
end
for j=1:length(MatComm)-1
test=abs((MatComm(1,j+1)-MatComm(1,j))/sqrt(MatComm(1,j)));
if test>=par*sqrt(sigma2)
M1=[M1, MatComm(2,j)];
M2 = [M2,j];
end
end
jumptimes=[jumptimes,M1];
m1=length(M1);
numberjumpstep=[numberjumpstep,m1];
sigmavec=[sigmavec,sqrt(sigma2)];
if k==3
m1=0;
end
MatComm(:, M2)=[];
end
```


[^0]:    ${ }^{1}$ Let $I$ denote an interval on the real line of the type $[a, b]$ and let $\beta, u$ two real functions continuous in the interior of $I$. If $u$ is differential in the interior of $I$ and satisfies $u^{\prime}(t) \leq \beta(t) u(t)$ then $u$ is bounded by the solution of the differential equation $\nu^{\prime}(t)=\beta(t) \nu(t)$, specifically $u(t) \leq u(a) \int_{0}^{t} \beta(s) d s$

[^1]:    ${ }^{1}$ GCM23.CMX on "https://finance.yahoo.com/", data are choose from 2017/6/29 to 2023/4/6
    ${ }^{2}$ CLN23.NYM on "https://finance.yahoo.com/"data are choose from 2017/11/21 to 2023/4/6

