# A splitting algorithm for stochastic partial differential equations driven by linear multiplicative noise 

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#### Abstract

We study the convergence of a Douglas-Rachford type splitting algorithm for the infinite dimensional stochastic differential equation $$
d X+A(t)(X) d t=X d W \text { in }(0, T) ; X(0)=x
$$ where $A(t): V \rightarrow V^{\prime}$ is a nonlinear, monotone, coercive and demicontinuous operator with sublinear growth and $V$ is a real Hilbert space with the dual $V^{\prime} . V$ is densely and continuously embedded in the Hilbert space $H$ and $W$ is an $H$-valued Wiener process. The general case of a maximal monotone operators $A(t): H \rightarrow H$ is also investigated.


Keywords: Maximal monotone operator, stochastic process, parabolic stochastic equation.

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Secondary 47H05, 47J05.

## 1 Introduction

We consider here the stochastic differential equation

$$
\begin{align*}
& d X(t)+A(t) X(t) d t=X(t) d W(t), t \in(0, T),  \tag{1.1}\\
& X(0)=x,
\end{align*}
$$

[^0]in a real separable Hilbert space $H$, whose elements are functions or distributions on a bounded and open set $\mathcal{O} \subset \mathbb{R}^{d}$ with smooth boundary $\partial \mathcal{O}$. In particular, $H$ can be any of the spaces $L^{2}(\mathcal{O}), H_{0}^{1}(\mathcal{O}), H^{-1}(\mathcal{O}), H^{1}(\mathcal{O})$, $k=1,2, \ldots$, with the corresponding Hilbertian structure. Here $H_{0}^{1}(\mathcal{O}), H^{k}(\mathcal{O})$ are the standard $L^{2}$-Sobolev spaces on $\mathcal{O}$, and $W$ is a Wiener process of the form
\[

$$
\begin{equation*}
W(t, \xi)=\sum_{j=1}^{\infty} \mu_{j} e_{j}(\xi) \beta_{j}(t), \xi \in \mathcal{O}, t \geq 0 \tag{1.2}
\end{equation*}
$$

\]

where $\left\{\beta_{j}\right\}_{j=1}^{\infty}$ is an independent system of real-valued Brownian motions on a probability space $\{\Omega, \mathcal{F}, \mathbb{P}\}$ with natural filtration $\left(\mathcal{F}_{t}\right)_{t \geq 0}$. Here, $e_{j} \in$ $C^{2}(\overline{\mathcal{O}}) \cap H, j \in \mathbb{N}$, is an orthonormal basis in $H$, and $\mu_{j} \in \mathbb{R}, j=1,2, \ldots$.

The following hypotheses will be in effect throughout this work.
(i) There is a Hilbert space $V$ with dual $V^{\prime}$ such that $V \subset H$, continuously and densely. Hence $V \subset H\left(\equiv H^{\prime}\right) \subset V^{\prime}$ continuously and densely.
(ii) $A:[0, T] \times V \times \Omega \rightarrow V^{\prime}$ is progressively measurable, i.e., for every $t \in[0, T]$, this operator restricted to $[0, t] \times V \times \Omega$ is $\mathcal{B}([0, t]) \otimes \mathcal{B}(V) \otimes \mathcal{F}_{t^{-}}$ measurable.
(iii) There is $\delta \geq 0$ such that, for each $t \in[0, T], \omega \in \Omega$, the operator $u \mapsto \delta u+A(t, \omega) u$ is monotone and demicontinuous (that is, stronglyweakly continuous) from $V$ to $V^{\prime}$.
Moreover, there are $\alpha_{i}, \gamma_{i} \in \mathbb{R}, i=1,2,3, \alpha_{1}>0$, such that, $\mathbb{P}$-a.s.,

$$
\begin{align*}
\langle A(t, \omega) u, u\rangle & \geq \alpha_{1}|u|_{V}^{2}+\alpha_{2}|u|_{H}^{2}+\alpha_{3}, & \forall u \in V, t \in[0, T]  \tag{1.3}\\
|A(t, \omega) u|_{V^{\prime}} & \leq \gamma_{1}|u|_{V}+\gamma_{2}, & \forall u \in V, t \in[0, T] . \tag{1.4}
\end{align*}
$$

(iv) $e^{ \pm W(t)}$ is, for each $t$, a multiplier in $V$ and a multiplier in $H$ such that there exists an $\left(\mathcal{F}_{t}\right)$-adapted, $\mathbb{R}_{+}$-valued process $Z(t), t \in[0, T]$, with $\mathbb{E}\left[\sup _{t \in[0, T]}|Z(t)|\right]<\infty$ for all $r \in[1, \infty)$ and such that, $\mathbb{P}$-a.s.,

$$
\begin{align*}
& \left|e^{ \pm W(t)} y\right|_{V} \leq Z(t)|y|_{V}, \quad \forall t \in[0, T], \quad \forall y \in V \\
& \left|e^{ \pm W(t)} y\right|_{H} \leq Z(t)|y|_{H}, \quad \forall t \in[0, T], \quad \forall y \in H \tag{1.5}
\end{align*}
$$

One assumes also that, for each $\omega \in \Omega$, the function $t \rightarrow e^{ \pm W(t)}$ is $H$-valued continuous on $[0, T]$.

Throughout in the following, $|\cdot|_{V}$ and $|\cdot|_{V^{\prime}}$ denote the norms of $V$ and $V^{\prime}$, respectively, and by $\langle\cdot, \cdot\rangle$ we denote the duality pairing between $V$ and $V^{\prime}$ with $H$ as pivot space; on $H \times H,\langle\cdot, \cdot\rangle$ is just the scalar product of $H$. The norms of $V$ and $V^{\prime}$ are denoted by $|\cdot|_{H}$ and $|\cdot|_{V},|\cdot|_{V^{\prime}}$, respectively, $\mathcal{B}(H), \mathcal{B}(V)$ etc. are the classes of Borel sets in the corresponding spaces.

As regards the orthonormal basis $\left\{e_{j}\right\}_{j=1}^{\infty}$ in (1.2), we assume that there exist $\widetilde{\gamma}_{j} \in[1, \infty)$ such that

$$
\begin{equation*}
\left|y e_{j}\right|_{H} \leq \widetilde{\gamma}_{j}|y|_{H}, \quad \forall y \in H, j=1,2, \ldots, \nu:=\sum_{j=1}^{\infty} \mu_{j}^{2} \widetilde{\gamma}_{j}^{2}\left|e_{j}\right|_{\infty}^{2}<\infty \tag{1.6}
\end{equation*}
$$

and we assume also that

$$
\begin{equation*}
\mu:=\frac{1}{2} \sum_{j=1}^{\infty} \mu_{j}^{2} e_{j}^{2} \tag{1.7}
\end{equation*}
$$

is a multiplier in $V, V^{\prime}$ and $H$.
It should be noted that $X d W=\sigma(X) d \widetilde{W}$ where $\sigma: H \rightarrow L_{2}(H)$ (the space of Hilbert-Schmidt operators on $H$ ) is defined by

$$
\sigma(u) v=\sum_{j=1}^{\infty} \mu_{j} u\left\langle v, e_{j}\right\rangle e_{j}, \quad \forall v \in H
$$

and so, $\widetilde{W}=\sum_{j=1}^{\infty} e_{j} \beta_{j}$ is a cylindrical Wiener process on $H$ (see [5]).
Definition 1.1. By a solution to (1.1) for $x \in H$, we mean an $\left(\mathcal{F}_{t}\right)_{t \geq 0^{-}}$ adapted process $X:[0, T] \rightarrow H$ with continuous sample paths which satisfies

$$
\begin{gather*}
X(t)+\int_{0}^{t} A(s) X(s) d s=x+\int_{0}^{t} X(s) d W(s), t \in[0, T]  \tag{1.8}\\
X \in L^{2}((0, T) \times \Omega ; V) \tag{1.9}
\end{gather*}
$$

The stochastic integral arising in (1.8) is considered in Itô's sense.
In [3], the authors developed an operatorial approach to (1.1) under the more general hypotheses than (i)-(iv) above. As a special case (see Theorem 3.1 in [3]), we have

Theorem 1.2. Under Hypotheses (i)-(iv), for each $x \in H$, equation (1.1) has a unique solution $X$ (in the sense of Definition 1.1). Moreover, the function $t \mapsto e^{-W(t)} X(t)$ is $V^{\prime}$-absolutely continuous on $[0, T]$ and

$$
\begin{equation*}
\mathbb{E} \int_{0}^{T}\left|e^{W(t)} \frac{d}{d t}\left(e^{-W(t)} X(t)\right)\right|_{V^{\prime}}^{2} d t<\infty \tag{1.10}
\end{equation*}
$$

In a few words, the method developed in [3] is the following. By the transformation

$$
\begin{equation*}
X(t)=e^{W(t)} y(t), t \geq 0 \tag{1.11}
\end{equation*}
$$

one reduces equation (1.1) to the random differential equation

$$
\begin{align*}
& \frac{d y}{d t}(t)+e^{-W(t)} A(t)\left(e^{W(t)} y(t)\right)+\mu y(t)=0, \text { a.e. } t \in(0, T)  \tag{1.12}\\
& y(0)=x
\end{align*}
$$

and treat (1.12) as an operatorial equation of the form

$$
\begin{equation*}
\mathcal{B} y+\mathcal{A} y=0 \tag{1.13}
\end{equation*}
$$

in a suitable Hilbert space $\mathcal{H}$ of stochastic processes on $[0, T]$. Here, $\mathcal{A}$ and $\mathcal{B}$ are maximal monotone operators suitable defined from $\mathcal{V}$ to $\mathcal{V}^{\prime}$, where $\left(\mathcal{V}, \mathcal{V}^{\prime}\right)$ is a dual pair of spaces such that $\mathcal{V} \subset \mathcal{H} \subset \mathcal{V}^{\prime}$ with dense and continuous embeddings.

The operatorial form (1.13) of equation (1.12) suggests to approximate the solution $y$ by the Douglas-Rachford splitting algorithm ([6]-[8]).

The exact form and convergence of the corresponding splitting algorithm for equation (1.13) will be given below in Section 2. As seen later on in Theorem 2.1, it leads to a convergent splitting algorithm for the stochastic differential equation (1.1).

In this way, the operator theoretic approach to equation (1.1) written in the form (1.13) allows to design a convergent splitting scheme for equation (1.1) inspired by the Rockafellar [9] proximal point algorithm for nonlinear operatorial equations (on these lines see also [4]). By our knowledge, the splitting algorithm obtained here for the stochastic equation is new and might have implications in numerical approximation of stochastic PDEs.

Notations. If $U$ is a Banach space, we denote by $L^{p}(0, T ; U), 1 \leq p \leq \infty$, the space of all $L^{p}$-integrable $U$-valued functions on $(0, T)$. The space
$L^{p}((0, T) \times \Omega ; U)$ is defined similarly. We refer to [2] for notation and standard results of the theory of maximal monotone operators in Banach spaces. If $\mathcal{O}$ is an open domain of $\mathbb{R}^{d}$, we denote by $W^{1, p}(\mathcal{O}), 1 \leq p \leq \infty$ and $H^{1}(\mathcal{O}), H^{-1}(\mathcal{O})$ the standard Sobolev spaces on $\mathcal{O}$.

## 2 Main results

Without loss of generality, we may assume that, besides assumptions (i)-(iii), $A(t)$ satisfies also the strong monotonicity condition

$$
\begin{equation*}
\langle A(t) u-A(t) v, u-v\rangle \geq \nu|u-v|_{H}^{2}, \forall u, v \in V, \tag{2.1}
\end{equation*}
$$

where $\nu>0$ is given by (1.6). (In fact, as easily seen, by the substitution $X \rightarrow \exp (-(\nu+\delta) t) X$ with a suitable $\delta$, equation (1.1) can be rewritten as

$$
d X+\widetilde{A}(t) X d t=X d W
$$

where the operator $X \rightarrow \widetilde{A}(t) X=e^{-(\nu+\delta) t} A(t)\left(e^{(\nu+\delta) t} X\right)+(\nu+\delta) X$ satisfies conditions (i)-(iii) and (2.1).)

We associate with equation (1.1) the following splitting algorithm

$$
\begin{gather*}
\lambda d Z_{n+1}+J\left(Z_{n+1}\right) d t+\lambda \nu Z_{n+1} d t=\lambda Z_{n+1} d W-\lambda A(t) X_{n} d t \\
\quad+\lambda \nu X_{n} d t+J\left(X_{n}\right) d t, t \in(0, T),  \tag{2.2}\\
Z_{n+1}(0)=x, n=0,1, \ldots \\
\lambda A(t) X_{n+1}(t)+J\left(X_{n+1}(t)\right)-\lambda \nu X_{n+1}(t) \\
=J\left(Z_{n+1}(t)\right)+\lambda A(t) X_{n}(t)-\lambda \nu X_{n}(t), \tag{2.3}
\end{gather*}
$$

where $X_{0} \in L^{2}((0, T) \times \Omega ; V)$ is $\left(\mathcal{F}_{t}\right)_{t \geq 0}$-adapted and arbitrary. Here, the parameter $\lambda>0$ is arbitrary but fixed and $J: V \rightarrow V^{\prime}$ is the canonical isomorphism of the space $V$ onto its dual $V^{\prime}$.

Taking into account assumptions (i)-(iii) and (2.1), which, in particular, implies that the operator $\Gamma_{0}: L^{2}(0, T ; V) \rightarrow L^{2}\left(0, T ; V^{\prime}\right), \Gamma_{0} u=\lambda A(t) u+$ $J(u)-\lambda \nu u, u \in L^{2}(0, T ; V)$, is demicontinuous, locally bounded, and with inverse continuous, we see that the sequence $\left(Z_{n}, X_{n}\right)$ is well defined by (2.2), (2.3) and we have also

$$
\begin{equation*}
X_{n}, Z_{n} \in L^{2}((0, T) \times \Omega ; V) \text { and } Z_{n} \in L^{2}(\Omega ; C([0, T] ; H)), n=1,2, \ldots \tag{2.4}
\end{equation*}
$$

Moreover, the processes $X_{n}, Z_{n}$ are $\left(\mathcal{F}_{t}\right)_{t \geq 0}$-adapted on $[0, T]$.
Theorem 2.1 is the main result.

Theorem 2.1. Under Hypotheses (i)-(iv) and (2.1), assume that $x \in V$ and $\lambda>0$. If $\left(X_{n}, Z_{n}\right)$ is the sequence defined by (2.2), (2.3), we have for $n \rightarrow \infty$

$$
\begin{equation*}
X_{n} \rightarrow X \text { weakly in } L^{2}((0, T) \times \Omega ; V), \tag{2.5}
\end{equation*}
$$

where $X$ is the solution to equation (1.1) given by Theorem 1.2. Assume further that the operator $u \rightarrow A(t) u$ is odd, that is, $A(t)(-u)=-A(t) u$, $\forall u \in V$. Then, for $n \rightarrow \infty$,

$$
\begin{equation*}
X_{n} \rightarrow X \text { strongly in } L^{2}((0, T) \times \Omega ; V) \tag{2.6}
\end{equation*}
$$

The splitting scheme (2.2)-(2.3) reduces the approximation of problem (1.1) to a sequence of simpler linear equations. In fact, at each step $n$, one should solve a linear stochastic differential equation of the form

$$
\begin{align*}
& d Z_{n+1}+\frac{1}{\lambda} J\left(Z_{n+1}\right) d t+\nu Z_{n+1} d t=Z_{n+1} d W+F_{n} d t, t \in(0, T)  \tag{2.7}\\
& Z_{n+1}(0)=x
\end{align*}
$$

and the stationary random equation (2.3), where

$$
F_{n}=-\lambda A(t) X_{n}+\lambda \nu X_{n}+J\left(X_{n}\right) .
$$

By Itô's formula (see, e.g., [3]), equation (2.7) has, for each $n$, the solution

$$
Z_{n+1}=e^{W} z_{n+1}
$$

where $z_{n+1}$ is the solution to the random differential equation

$$
\begin{align*}
& \frac{d}{d t} z_{n+1}+\frac{1}{\lambda} e^{-W} J\left(e^{W} z_{n+1}\right)+(\mu+\nu) z_{n+1}=e^{-W} F_{n},  \tag{2.8}\\
& z_{n+1}(0)=x
\end{align*}
$$

If $F: L^{2}\left((0, T) \times \Omega ; V^{\prime}\right) \rightarrow L^{2}((0, T) \times \Omega ; V)$ is the linear continuous operator defined by

$$
F(f)=Y,
$$

where $Y$ is the solution to the stochastic equation

$$
d Y+\frac{1}{\lambda} J(Y) d t+\nu Y d t=Y d W+f d t ; Y(0)=x
$$

then we may rewrite (2.2)-(2.3) as
$X_{n+1}=(\lambda(A-\nu I)+J)^{-1}\left[J F\left((\lambda(\nu I-A)+J)\left(X_{n}\right)\right)+\lambda(A-\nu I) X_{n}\right], n=0,1, \ldots$
Equivalently,

$$
\begin{equation*}
X_{n+1}=\Gamma^{n} X_{0}, \forall n \in \mathbb{N}, \tag{2.9}
\end{equation*}
$$

where $\Gamma: L^{2}((0, T) \times \Omega ; V) \rightarrow L^{2}((0, T) \times \Omega ; V)$ is the Lipschitzian and given by

$$
\begin{equation*}
\Gamma=(\lambda(A-\nu I)+J)^{-1}[J F(\lambda(\nu I-A)+J)+\lambda(A-\nu I)] . \tag{2.10}
\end{equation*}
$$

Then, by Theorem 2.1, we get
Corollary 2.2. Under assumptions (i)-(iv), (2.1), for each $\lambda>0$ the solution $X$ to (1.1) is expressed as

$$
\begin{equation*}
X=w-\lim _{n \rightarrow \infty} \Gamma^{n} X_{0} \quad \text { in } L^{2}((0, T) \times \Omega ; V) \tag{2.11}
\end{equation*}
$$

where $X_{0} \in L^{2}((0, T) \times \Omega ; V)$ is an arbitrary $\left(\mathcal{F}_{t}\right)_{t \geq 0}$-adapted process.
Here $w-\lim$ indicates the weak limit.

## 3 Proof of Theorem 2.1

Proceeding as in [3], we consider the spaces $\mathcal{H}, \mathcal{V}$ and $\mathcal{V}^{\prime}$, defined as follows. $\mathcal{H}$ is the Hilbert space of all $\left(\mathcal{F}_{t}\right)_{t \geq 0}$-adapted processes $y:[0, T] \rightarrow H$ such that

$$
|y|_{\mathcal{H}}=\left(\mathbb{E} \int_{0}^{Y}\left|e^{W(t)} y(t)\right|_{H}^{2} d t\right)^{\frac{1}{2}}<\infty
$$

where $\mathbb{E}$ denotes the expectation in the probability space $(\Omega, \mathcal{F}, \mathbb{P})$. The space $\mathcal{H}$ is endowed with the norm $|\cdot|_{\mathcal{H}}$ generated by the scalar product

$$
\langle y, z\rangle_{H}=\mathbb{E} \int_{0}^{T}\left\langle e^{W(t)} y(t), e^{E(t)} y(t)\right\rangle d t
$$

$\mathcal{V}$ is the space of all $\left(\mathcal{F}_{t}\right)_{t \geq 0}$-adapted processes $y:[0, T] \rightarrow V$ such that

$$
|y|_{\mathcal{V}}=\left(\mathbb{E} \int_{0}^{T}\left|e^{W(t)} y(t)\right|_{V}^{2} d t\right)^{\frac{1}{2}}<\infty
$$

$\mathcal{V}^{\prime}$ (the dual of $\left.\mathcal{V}\right)$ is the space of all $\left(\mathcal{F}_{t}\right)_{t \geq 0}$-adapted processes $y:[0, T] \rightarrow V^{\prime}$ such that

$$
|y|_{\mathcal{V}^{\prime}}=\left(\mathbb{E} \int_{0}^{T}\left|e^{W(t)} y(t)\right|_{V^{\prime}}^{2} d t\right)^{\frac{1}{2}}<\infty
$$

We have $\mathcal{V} \subset \mathcal{H} \subset \mathcal{V}^{\prime}$ with continuous and dense embeddings. Moreover,

$$
\mathcal{V}^{\prime}\langle u, v\rangle_{\mathcal{V}}=\mathbb{E} \int_{0}^{T}\left\langle e^{W(t)} u(t), e^{W(t)} v(t)\right\rangle d t, v \in \mathcal{V}, u \in \mathcal{V}^{\prime}
$$

is the duality pairing between $\mathcal{V}$ and $\mathcal{V}^{\prime}$, with the pivot space $\mathcal{H}$, that is,

$$
\mathcal{V}^{\prime}\langle u, v\rangle_{\mathcal{V}}=\langle u, v\rangle_{\mathcal{H}}, \forall u \in \mathcal{H}, v \in \mathcal{V}
$$

Now, for $x \in H$, define the operators $\mathcal{A}: \mathcal{V} \rightarrow \mathcal{V}^{\prime}$ and $\mathcal{B}: D(\mathcal{B}) \subset \mathcal{V} \rightarrow \mathcal{V}^{\prime}$ as follows:

$$
\begin{align*}
&(\mathcal{A} y)(t)=e^{-W(t)} A(t)\left(e^{W(t)} y(t)\right)-\nu y(t), \text { a.e. } t \in(0, T), y \in \mathcal{V} \\
&(\mathcal{B} y)(t)=\frac{d y}{d t}(t)+(\mu+\nu) y(t), \text { a.e. } t \in(0, T), y \in D(\mathcal{B})  \tag{3.1}\\
& D(\mathcal{B})=\left\{y \in \mathcal{V}: y \in A C\left([0, T] ; V^{\prime}\right) \cap C([0, T] ; H), \mathbb{P}\right. \text {-a.s. } \\
&\left.\frac{d y}{d t} \in \mathcal{V}^{\prime}, y(0)=x\right\} \tag{3.2}
\end{align*}
$$

Here, $A C\left([0, T] ; V^{\prime}\right)$ is the space of all absolutely continuous $V^{\prime}$-valued functions on $[0, T]$. If $y \in D(\mathcal{B})$, then $y \in C([0, T] ; H)$ and $\frac{d y}{d t}$ is the derivative of $y$ in the sense of $V^{\prime}$-valued distributions on $(0, T)$. Then, equation (1.12) can be expressed as

$$
\begin{equation*}
\mathcal{B} y+\mathcal{A} y=0 \tag{3.3}
\end{equation*}
$$

Then, the map $\Lambda: \mathcal{V} \rightarrow \mathcal{V}^{\prime}$ defined by

$$
\begin{equation*}
\Lambda v=e^{-W} J\left(e^{W} v\right), v \in V \tag{3.4}
\end{equation*}
$$

is the canonical isomorphism of $\mathcal{V}$ onto $\mathcal{V}^{\prime}$ and the scalar product $\mathcal{V}\langle\cdot, \cdot\rangle_{\mathcal{V}}$ of the space $\mathcal{V}$ can be expressed as

$$
\begin{equation*}
\mathcal{\nu}\langle v, \bar{v}\rangle_{\mathcal{V}}=\mathcal{V}\langle v, \Lambda \bar{v}\rangle_{\mathcal{V}^{\prime}}, \forall v, \bar{v} \in \mathcal{V} \tag{3.5}
\end{equation*}
$$

We set

$$
\begin{array}{r}
\left(\mathcal{A}^{*} u\right)(t)=\Lambda^{-1} \mathcal{A} u(t)=e^{-W} J^{-1}\left(A(t)\left(e^{W} u\right)-\nu e^{W} u\right), \forall u \in \mathcal{V}, \\
\left(\mathcal{B}^{*} u\right)(t)=\Lambda^{-1} \mathcal{B} u(t)=e^{-W} J^{-1}\left(e^{W}\left(\frac{d u}{d t}+(\mu+\nu) u\right)\right),  \tag{3.7}\\
\forall u \in D\left(\mathcal{B}^{*}\right)=D(\mathcal{B}) .
\end{array}
$$

Since the operators $\mathcal{A}, \mathcal{B}$ and $\mathcal{A}+\mathcal{B}$ are maximal monotone in $\mathcal{V} \times \mathcal{V}^{\prime}([3]$, Lemma 4.1, Lemma 4.2), it is easily seen by (3.6)-(3.7) that $\mathcal{A}^{*}, \mathcal{B}^{*}$ and $\mathcal{A}^{*}+\mathcal{B}^{*}$ are maximal monotone in $\mathcal{V} \times \mathcal{V}$.

On the other hand, by (3.3) we can rewrite equation (3.3) as

$$
\begin{equation*}
\mathcal{B}^{*} y+\mathcal{A}^{*} y=0 \tag{3.8}
\end{equation*}
$$

Let $y \in D(\mathcal{B})$ be the unique solution to equation (3.3) (see [3], Proposition 3.3). Then, $y$ is also the solution to (3.8) and so, by Theorem 1 in [8] (see, also, Corollary 6.1 in [7]), we have that

$$
\begin{equation*}
y=\lim _{n \rightarrow \infty}\left(I+\lambda \mathcal{A}^{*}\right)^{-1} v_{n} \text { weakly in } \mathcal{V} \text { as } n \rightarrow \infty \tag{3.9}
\end{equation*}
$$

where $\left\{v_{n}\right\} \subset \mathcal{V}$ is, for $n \geq 0$, defined by

$$
\begin{equation*}
v_{n+1}=\left(I+\lambda \mathcal{B}^{*}\right)^{-1}\left(2\left(I+\lambda \mathcal{A}^{*}\right)^{-1} v_{n}-v_{n}\right)+\left(I-\left(I+\lambda \mathcal{A}^{*}\right)^{-1}\right) v_{n} \tag{3.10}
\end{equation*}
$$

and $v_{0}$ is arbitrary in $\mathcal{V}$. Here, $I$ is the identity operator in $\mathcal{V}$.
The splitting algorithm (3.9)-(3.10) is just the Douglas-Rachford algorithm ([6]) for equation (3.8) and it can be equivalently expressed as

$$
\begin{align*}
y & =\lim _{n \rightarrow \infty} y_{n} \text { weakly in } \mathcal{V},  \tag{3.11}\\
y_{n} & =\left(I+\lambda \mathcal{A}^{*}\right)^{-1} v_{n}, n=0,1, \ldots,  \tag{3.12}\\
y_{n+1}+\lambda \mathcal{A}^{*} y_{n+1} & =z_{n+1}+v_{n}-y_{n},  \tag{3.13}\\
z_{n+1}+\lambda \mathcal{B}^{*} z_{n+1} & =2 y_{n}-v_{n}, \tag{3.14}
\end{align*}
$$

where $v_{0} \in \mathcal{V}$. (To get (3.12)-(3.14) from (3.10), we have used the identity $\left(I+\lambda \mathcal{B}^{*}\right)^{-1}\left(v+\lambda \mathcal{B}^{*} v\right)=v, \forall v \in D\left(\mathcal{B}^{*}\right)$ and the linearity of $\mathcal{B}^{*}$.)

In fact, the weak convergence of $\left\{v_{n}\right\}$ in the space $\mathcal{V}$ is also a consequence of the convergence of the Rockafellar proximal point algorithm [9] for the maximal monotone operator $v \rightarrow G^{-1}(v)-v$, where

$$
\begin{equation*}
G(z)=\left(I+\lambda \mathcal{B}^{*}\right)^{-1}\left(2\left(I+\lambda \mathcal{A}^{*}\right)^{-1} z-z\right)+z-\left(I+\lambda \mathcal{A}^{*}\right)^{-1} z, \forall z \in \mathcal{V} \tag{3.15}
\end{equation*}
$$

(See [7], Theorem 4.) Taking into account (3.6), (3.7), (3.12) we rewrite (3.14) as

$$
\begin{array}{rl}
e^{-W} & J\left(e^{W} z_{n+1}\right)+\lambda\left(\frac{d z_{n+1}}{d t}+(\mu+\nu) z_{n+1}\right) \\
& =e^{-W} J\left(e^{W}\left(2 y_{n}-v_{n}\right)\right)=e^{-W} J\left(e^{W}\left(-\lambda \mathcal{A}^{*} y_{n}+y_{n}\right)\right)  \tag{3.16}\\
& =-\lambda e^{-W} A(t)\left(e^{W} y_{n}\right)+\lambda \nu y_{n}+e^{-W} J\left(e^{W} y_{n}\right)
\end{array}
$$

and (3.13) as

$$
\begin{array}{r}
J\left(e^{W} y_{n+1}\right)+\lambda A(t)\left(e^{W} y_{n+1}\right)-\lambda \nu e^{W} y_{n+1}  \tag{3.17}\\
=J\left(e^{W}\left(z_{n+1}+v_{n}-y_{n}\right)\right) .
\end{array}
$$

We set

$$
X_{n}=e^{W} y_{n}, Z_{n}=e^{W} z_{n} .
$$

Then, by (3.16), we get via Itô's formula (see [3] and (2.8), (2.7))

$$
\begin{array}{lr}
\lambda d Z_{n+1}+J\left(Z_{n+1}\right) d t+\lambda \nu Z_{n+1} d t=\lambda Z_{n+1} d W-\lambda A(t) X_{n} d t \\
& +\lambda \nu X_{n} d t+J\left(X_{n}\right) d t \\
Z_{n+1}(0)=x . &
\end{array}
$$

By (3.17) and (3.12), we also get that

$$
\begin{aligned}
& \lambda A(t) X_{n+1}(t)+J\left(X_{n+1}(t)\right)-\lambda \nu X_{n+1}(t) \\
& \quad=J\left(Z_{n+1}(t)\right)+\lambda A(t) X_{n}(t)-\lambda \nu X_{n}(t), t \in(0, T)
\end{aligned}
$$

which are just equations (2.2), (2.3). Moreover, by (3.11), we see that (2.5) holds.

Assume now that $A(t): V \rightarrow V^{\prime}$ is odd. Then so is $\mathcal{A}^{*}: \mathcal{V} \rightarrow \mathcal{V}$ and also the operator $G$ defined by (3.15). Then, according to a result of J. Baillon [1], the sequence $\left\{v_{n}\right\}$ defined by (3.10), that is $v_{n+1}=G\left(v_{n}\right)$, is strongly convergent in $\mathcal{V}$. Recalling (3.9), we infer that so is the sequence $\left\{y_{n}\right\}$ and, consequently, (2.6) holds. This completes the proof of Theorem 2.1.

Remark 3.1. One might expect that a similar splitting scheme can be constructed for nonlinear monotone operators $A(t): V \rightarrow V^{\prime}$, where $V$ is a
reflexive Banach space and $A(t)$ are demicontinuous coercive and with polynomial growth as in [3]. In fact, in this case, one might replace (2.2) by

$$
\begin{aligned}
& \lambda d Z_{n+1}+Z_{n+1} d t+\lambda \nu Z_{n+1} d t=Z_{n+1} d W-\lambda A_{H}(t) X_{n} d t+\lambda \nu X_{n} d t+X_{n} d t, \\
& \quad t \in(0, T), \\
& \lambda A_{H}(t) X_{n+1}+X_{n+1}-\lambda \nu X_{n+1}=Z_{n+1}+\lambda A_{H}(t) X_{n}-\lambda \nu X_{n},
\end{aligned}
$$

where $A_{H}(t) u=A(t) u \cap H$. This question will be addressed in Section 5 below (see Remark 5.2).

## 4 Examples

We shall illustrate here the splitting algorithm (2.2)-(2.3) for a few parabolic stochastic differential equations.

Example 4.1. Nonlinear stochastic parabolic equations.
Consider the reaction-diffusion stochastic equation in $\mathcal{O} \subset \mathbb{R}^{d}$,

$$
\begin{align*}
& d X-\operatorname{div}(a(t, \xi, \nabla X)) d t+\nu X d t+\psi(X) d t=X d W \text { in }(0, T) \times \mathcal{O}  \tag{4.1}\\
& X=0 \text { on }(0, T) \times \partial \mathcal{O}, \quad X(0)=x \text { in } \mathcal{O}
\end{align*}
$$

Here, $a:(0, T) \times \mathcal{O} \times \mathbb{R}^{d} \rightarrow \mathbb{R}^{d}$ is measurable in $(t, \xi, r)$ continuous in $r$ on $\mathbb{R}^{d}$, $a(t, \xi, 0)=0$. (The more general case, when $\left.a:(0, T) \times \mathcal{O} \times \Omega \times \mathbb{R}^{d}\right) \rightarrow \mathbb{R}^{d}$ is progressively measurable, could also be considered.) We assume also that

$$
\begin{array}{ll}
\left(a\left(t, \xi, r_{1}\right)-a\left(t, \xi, r_{2}\right)\right) \cdot\left(r_{1}-r_{2}\right) \geq 0, & \forall r_{1}, r_{2} \in \mathbb{R}^{d},(t, \xi) \in(0, T) \times \mathcal{O}, \\
a(t, \xi, r) \cdot r \geq a_{1}|r|_{d}^{2}+a_{2}, & \forall r \in \mathbb{R}^{d},(t, \xi) \in(0, T) \times \mathcal{O} \\
|a(t, \xi, r)|_{d} \leq c_{1}|r|_{d}+c_{2}, & \forall r \in \mathbb{R}^{d},(t, \xi) \in(0, T) \times \mathcal{O}
\end{array}
$$

where $a_{1}, c_{1}, \nu>0, a_{2}, c_{2} \in \mathbb{R}$, are independent of $(t, \xi)$, and $\psi: \mathbb{R} \rightarrow \mathbb{R}$ is a continuous and monotonically nondecreasing function such that $\psi(0)=0$ and $|\psi(r)| \leq C\left(|r|^{\frac{2 d}{d+2}}+1\right), \forall r \in \mathbb{R}$. Here $\mathcal{O} \subset \mathbb{R}^{d}$ is a bounded open subset with smooth boundary $\partial \mathcal{O}$, and $|\cdot|_{d}$ is the Euclidean norm of $\mathbb{R}^{d}$.

If $H=L^{2}(\mathcal{O}), V=H_{0}^{1}(\mathcal{O}), V^{\prime}=H^{-1}(\mathcal{O})$ and, for $t \in(0, T)$, the operator $A(t): V \rightarrow V^{\prime}$ is defined by

$$
V^{\prime}\langle A(t) y, \varphi\rangle_{V}=\int_{\mathcal{O}}(a(t, \xi, \nabla y) \cdot \nabla \varphi+\psi(y) \varphi) d \xi, \forall \varphi \in H_{0}^{1}(\mathcal{O}), y \in H_{0}^{1}(\mathcal{O})
$$

then Hypotheses (i)-(iii) are satisfied. As regards the Wiener process $W$, we assume here that, besides (1.6), the following condition holds:

$$
\sum_{j=1}^{\infty} \mu_{j}^{2}\left|\nabla e_{j}\right|_{\infty}^{2}<\infty
$$

Then, by Theorem 2.1, where $H, V$ and $A(t)$ are defined above and $J=-\Delta$ with Dirichlet homogeneous boundary conditions, if $x \in H_{0}^{1}(\mathcal{O})$, the solution

$$
X \in L^{2}\left(\Omega ; C\left([0, T] ; L^{2}(\mathcal{O})\right) \cap L^{2}\left((0, T) \times \Omega ; H_{0}^{1}(\mathcal{O})\right)\right)
$$

to (4.1) can be obtained as

$$
\begin{equation*}
X=w-\lim _{n \rightarrow \infty} X_{n} \text { in } L^{2}\left((0, T) \times \Omega ; H_{0}^{1}(\mathcal{O})\right) \tag{4.2}
\end{equation*}
$$

where $\left(X_{n}, Z_{n}\right) \in L^{2}\left((0, T) \times \Omega ; H_{0}^{1}(\mathcal{O})\right)$ is the solution to the system (we take $\lambda=1$ )

$$
\begin{array}{r}
d Z_{n+1}-\Delta Z_{n+1} d t+\nu Z_{n+1} d t=Z_{n+1} d W+\operatorname{div}\left(a\left(t, \xi, \nabla X_{n}\right)\right) d t-\Delta X_{n} d t \\
\operatorname{in~}(0, T) \times \mathcal{O}
\end{array}
$$

$$
Z_{n+1}(0)=x \text { in } \mathcal{O},
$$

$$
Z_{n+1}=0 \quad \text { in }(0, T) \times \partial \mathcal{O},
$$

$$
\begin{equation*}
\operatorname{div} a\left(\nabla X_{n+1}\right)+\Delta X_{n+1}=\Delta Z_{n+1}+\operatorname{div}\left(a\left(t, \xi, \nabla X_{n}\right)\right) \text { in }(0, T) \times \mathcal{O} \tag{4.3}
\end{equation*}
$$

where $X_{0} \in L^{2}\left((0, T) \times \Omega ; H_{0}^{1}(\mathcal{O})\right)$ is arbitrary but $\mathcal{F}_{t}$-adapted. Moreover, if $a(t, \xi,-r) \equiv-a(t, \xi, r), \forall r \in \mathbb{R}^{d}$, then the convergence (4.2) is strong in $L^{2}\left((0, T) \times \Omega ; H_{0}^{1}(\mathcal{O})\right)$.

Example 4.2. Stochastic porous media equations.
Consider the stochastic equation

$$
\begin{array}{ll}
d X-\Delta \psi(t, \xi, X) d t-\nu \Delta X d t=X d W & \text { in }(0, T) \times \mathcal{O} \\
X(0, \xi)=x(\xi) & \text { in } \mathcal{O}  \tag{4.4}\\
\psi(t, \xi, X(t, \xi))=0 & \text { on }(0, T) \times \partial \mathcal{O}
\end{array}
$$

where $\mathcal{O}$ is a bounded domain in $\mathbb{R}^{d}, \nu>0$, the function $\psi:[0, T] \times \overline{\mathcal{O}} \times \mathbb{R} \rightarrow \mathbb{R}$ is continuous, $r \rightarrow \psi(t, \xi, r)$ is monotonically increasing in $r$, and there exist
$a \in(0, \infty)$ and $c \in[0, \infty)$ such that

$$
\begin{align*}
r \psi(t, \xi, r) & \geq a|r|^{2}-c, \quad \forall r \in \mathbb{R}, \quad(t, \xi, r) \in[0, T] \times \overline{\mathcal{O}}, \\
|\psi(t, \xi, r)| \leq c(1+|r|), & \forall r \in \mathbb{R}, \quad(t, \xi, r) \in[0, T] \times \overline{\mathcal{O}} \tag{4.5}
\end{align*}
$$

We shall write equation (4.4) under the form (1.1) with $H=H^{-1}(\mathcal{O})$. Namely, we take $V=L^{2}(\mathcal{O}), H=H^{-1}(\mathcal{O})$, and $V^{\prime}$ is the dual of $V$ with the pivot space $H^{-1}(\mathcal{O})$. Then, $V \subset H \subset V^{\prime}$ and

$$
V^{\prime}=\left\{\theta \in \mathcal{D}^{\prime}(\mathcal{O}): \theta=-\Delta v, v \in L^{2}(\mathcal{O})\right\}
$$

where $\Delta$ is taken in the sense of distributions on $\mathcal{O}$. (Here $\mathcal{D}^{\prime}(\mathcal{O})$ is the space of Schwartz distributions on $\mathcal{O}$.) The duality ${ }_{V^{\prime}}\langle\cdot, \cdot\rangle_{V}$ is defined as

$$
V_{V^{\prime}}\langle\theta, u\rangle_{V}=\int_{\mathcal{O}} \tilde{\theta} u d \xi, \quad \tilde{\theta}=(-\Delta)^{-1} \theta
$$

where $\Delta$ is the Laplace operator with homogeneous Dirichlet boundary conditions on $\partial \mathcal{O}$. The duality mapping $J: V \rightarrow V^{\prime}$ is just the operator $-\Delta$ defined from $L^{2}(\mathcal{O})$ to $V^{\prime} \subset \mathcal{D}^{\prime}(\mathcal{O})$ by

$$
\Delta u(\varphi)=\int_{\mathcal{O}} u \Delta \varphi d \xi, \forall \varphi \in H_{0}^{1}(\mathcal{O}) \cap H^{2}(\mathcal{O})
$$

The operator $A(t): V \rightarrow V^{\prime}$ is defined by

$$
V^{\prime}\langle A(t) y, v\rangle_{V}=\int_{\mathcal{O}} \psi(t, \xi, y) v d \xi, \quad \forall y, v \in V=L^{2}(\mathcal{O}), t \in[0, T]
$$

Then, Hypotheses (i)-(iv) hold and so, if $x \in L^{2}(\mathcal{O})$, by Theorem 2.1, the solution $X \in L^{2}\left(\Omega ; C\left([0, T] ; H^{-1}(\mathcal{O})\right) \cap L^{2}\left((0, T) \times \Omega ; L^{2}(\mathcal{O})\right)\right)$ to (4.4) is given by

$$
X=w-\lim _{n \rightarrow \infty} X_{n} \text { in } L^{2}\left((0, T) \times \Omega ; L^{2}(\mathcal{O})\right),
$$

where

$$
\begin{gather*}
d Z_{n+1}-\Delta Z_{n+1} d t+\nu Z_{n+1} d t=Z_{n+1} d W-\Delta \psi\left(t, \cdot, X_{n}\right) d t-\Delta Z_{n} d t \\
\quad \operatorname{} \quad \text { n }(0, T) \times \mathcal{O} \\
Z_{n+1}(0)=x \in L^{2}(\mathcal{O}), n=0,1, \ldots,  \tag{4.6}\\
\Delta \psi\left(t, \cdot, X_{n+1}\right)+X_{n+1}=Z_{n+1}+\Delta \psi\left(t, \cdot, X_{n}\right), \text { in } \mathcal{O}, \\
\psi\left(t, \cdot, X_{n+1}(t, \cdot)\right)=0 \text { on } \partial \mathcal{O}, \\
n=0,1, \ldots, X_{0} \in L^{2}\left(0, T ; L^{2}\left(\Omega ; L^{2}(\mathcal{O})\right)\right) .
\end{gather*}
$$

If $\psi(t, \xi, r)=-\psi(t, \xi,-r), \forall r \in \mathbb{R}$, then the convergence of the sequence $\left\{X_{n}\right\}$ is strong in $L^{2}\left((0, T) \times \Omega ; L^{2}(\mathcal{O})\right)$.

## 5 The case where $A(t)$ is maximal monotone in $H \times H$

Consider now equation (1.1) under the following assumptions on $A$ :
(j) $A:[0, T] \times H \times \Omega \rightarrow H$ is progressively measurable and, for each $(t, \omega) \times[0, T] \times \Omega$ the operator $u \rightarrow A(t, \omega, u)$ is maximal monotone in $H \times H$. Moreover, there is $f \in L^{2}((0, T) \times \Omega ; H)$ such that

$$
\begin{equation*}
(I+A(t))^{-1} f(t) \in L^{2}((0, T) \times \Omega ; H) \tag{5.1}
\end{equation*}
$$

We assume also that condition (2.1) holds.
It should be noted that, if $A(t): V \rightarrow V^{\prime}$ satisfies assumptions (i)-(ii), where $V$ is a reflexive Banach space, then the operator $A(t): H \rightarrow H$, defined by

$$
A(t)_{H} u=A(t) u \cap V,
$$

satisfies assumption (j). However, the class of the operators $A$ satisfying (j) is considerably larger.

We consider the splitting scheme (which is well defined by strong monotonicity of $\mathcal{A}_{1}^{*}+\mathcal{B}_{1}^{*}$ )

$$
\begin{align*}
\lambda d Y_{n+1} & +(1+\lambda \nu) Y_{n+1} d t=\lambda Y_{n+1} d W \\
& +\left(V_{n}-((1-\lambda \nu) I-\lambda A(t))^{-1} V_{n}\right) d t \\
Y_{n+1}(0) & =x \text { in }(0, T)  \tag{5.2}\\
V_{n+1}= & Y_{n+1}+V_{n}-((1-\lambda \nu) I-\lambda A(t))^{-1} V_{n}
\end{align*}
$$

where $V_{0} \in L^{2}((0, T) \times \Omega ; H)$ is an $\left(\mathcal{F}_{t}\right)_{t \geq 0}$-adapted process such that $A(t) V_{0} \in$ $L^{2}((0, T) \times \Omega ; H)$. We have

Theorem 5.1. Assume that $x \in H$ and that equation (1.1) has a solution $X \in L^{2}(\Omega ; C([0, T] ; H))$ such that $A(t) X \in L^{2}((0, T) \times \Omega ; H)$.

Then, for $n \rightarrow \infty$,

$$
\begin{equation*}
V_{n} \rightarrow V \text { weakly in } L^{2}((0, T) \times \Omega ; H), \tag{5.3}
\end{equation*}
$$

where $X=((1-\lambda \nu) I+A(t))^{-1} V$ is the solution to (1.1). If $A(t)$ is odd, then the convergence (5.3) is strong.

Proof. The operators $\mathcal{A}_{1}^{*}$ and $\mathcal{B}_{1}^{*}$ defined by

$$
\begin{aligned}
\left(\mathcal{A}_{1}^{*} u\right)(t) & =e^{-W} A(t)\left(e^{W} u\right)-\nu u, & & \forall u \in D\left(\mathcal{A}_{1}^{*}\right), \\
\left(\mathcal{B}_{1}^{*} u\right)(t) & =\frac{d u}{d t}+(\mu+\nu) u, & & \forall u \in D\left(\mathcal{B}_{1}^{*}\right),
\end{aligned}
$$

with the domains

$$
\begin{aligned}
D\left(\mathcal{A}_{1}^{*}\right) & =\left\{u \in \mathcal{H} ; e^{-W} A(t)\left(e^{W} u\right)-\nu u \in \mathcal{H}\right\} \\
D\left(\mathcal{B}_{1}^{*}\right) & =\left\{u \in \mathcal{H} ; u \in W^{1,2}([0, T] ; H) \cdot \mathbb{P} \text {-a.s., } u(0)=x\right\}
\end{aligned}
$$

are, by the above hypotheses, maximal monotone in $\mathcal{H} \times \mathcal{H}$ (see also [4]). Moreover, there is at least one solution $y^{*}$ to the equation

$$
\begin{equation*}
\mathcal{A}_{1}^{*} y^{*}+\mathcal{B}_{1}^{*} y^{*}=0 \tag{5.4}
\end{equation*}
$$

Then, again by [8], it follows that the sequence $\left\{v_{n}\right\} \subset \mathcal{H}$ defined by

$$
\begin{array}{r}
v_{n+1}=\left(I+\lambda \mathcal{B}_{1}^{*}\right)^{-1}\left(2\left(I+\lambda \mathcal{A}_{1}^{*}\right)^{-1} v_{n}-v_{n}\right)+v_{n}-\left(I+\lambda \mathcal{A}_{1}^{*}\right)^{-1} v_{n}  \tag{5.5}\\
n=0,1, \ldots
\end{array}
$$

is weakly convergent in $\mathcal{H}$ to $v^{*}$, where $\left(1+\lambda \mathcal{A}_{1}^{*}\right)^{-1} v^{*}=y^{*}$ is the solution to equation (5.4).

We set

$$
\begin{equation*}
\widetilde{z}_{n+1}=v_{n+1}-v_{n}+\left(I+\lambda \mathcal{A}_{1}^{*}\right)^{-1} v_{n} \tag{5.6}
\end{equation*}
$$

and, by (5.5), we have

$$
\begin{equation*}
\widetilde{z}_{n+1}+\lambda \mathcal{B}_{1}^{*} z_{n+1}=v_{n}-\left(I+\lambda \mathcal{A}_{1}^{*}\right)^{-1} v_{n} . \tag{5.7}
\end{equation*}
$$

Then, if $Y_{n}=e^{W} \widetilde{z}_{n}$ and $V_{n}=e^{W} v_{n}$, we can rewrite (5.6)-(5.7) as (5.2) and get (5.3), as claimed.

Remark 5.2. The convergence of the splitting algorithm (5.1)-(5.2) does not require conditions of the form (ii)-(iii) for the operator $A(t)$ but in change it requires the existence of a sufficiently regular solution $X$ for equation (1.1)
$\left(A(t) x \in L^{2}((0, T) \times \Omega ; H)\right)$ which is not the case for Examples 4.1, 4.2. Such a condition holds, however, for the stochastic reaction-diffusion equation

$$
\begin{aligned}
& d X-\Delta X d t+\Psi(X) d t=X d W \text { in }(0, T) \times \mathcal{O} \\
& X=0 \text { on }(0, T) \times \partial \mathcal{O} \\
& X(0)=x
\end{aligned}
$$

if $x \in H_{0}^{1}(\mathcal{O})$ and $\Psi: \mathbb{R} \rightarrow \mathbb{R}$ is continuous and monotonically increasing and for other stochastic parabolic equations as well.

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