# SOBOLEV REGULARITY FOR A CLASS OF SECOND ORDER ELLIPTIC PDE'S IN INFINITE DIMENSION 

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#### Abstract

We consider an elliptic Kolmogorov equation $\lambda u-K u=f$ in a separable Hilbert space $H$. The Kolmogorov operator $K$ is associated to an infinite dimensional convex gradient system: $d X=(A X-D U(X)) d t+d W(t)$, where $A$ is a self-adjoint operator in $H$, and $U$ is a convex lower semicontinuous function. Under mild assumptions we prove that for $\lambda>0$ and $f \in L^{2}(H, v)$ the weak solution $u$ belongs to the Sobolev space $W^{2,2}(H, v)$, where $v$ is the log-concave probability measure of the system. Moreover maximal estimates on the gradient of $u$ are proved. The maximal regularity results are used in the study of perturbed nongradient systems, for which we prove that there exists an invariant measure. The general results are applied to Kolmogorov equations associated to reaction-diffusion and Cahn-Hilliard stochastic PDEs.


1. Introduction. Let $H$ be an infinite dimensional separable Hilbert space (norm $\|\cdot\|$, inner product $\langle\cdot, \cdot\rangle$ ). We are concerned with the differential equation

$$
\begin{equation*}
\lambda u-\frac{1}{2} \operatorname{Tr}\left[D^{2} u\right]-\langle A x-D U(x), D u\rangle=f \tag{1.1}
\end{equation*}
$$

where $A: D(A) \subset H \rightarrow H$ is a linear self-adjoint negative operator, and such that $A^{-1}$ is of trace class, $U: H \rightarrow \mathbb{R} \cup\{+\infty\}$ is convex, proper, lowerly bounded, and lower semicontinuous. The data are $\lambda>0$ and $f: H \rightarrow \mathbb{R}$, the unknown is $u: H \rightarrow \mathbb{R} . D u$ and $D^{2} u$ represent first and second derivatives of $u$, and $\operatorname{Tr}\left[D^{2} u\right]$ is the trace of $D^{2} u$.

Equation (1.1) is the elliptic Kolmogorov equation corresponding to the differential stochastic equation

$$
\begin{align*}
d X & =(A X-D U(X)) d t+d W(t),  \tag{1.2}\\
X(0) & =x \tag{1.3}
\end{align*}
$$

where $W(t), t \geq 0$, is an $H$-valued cylindrical Wiener process. Equation (1.2) is a typical example of gradient system. Under suitable assumptions, it has a unique invariant measure $v(d x)=Z^{-1} e^{-2 U(x)} \mu(d x)$, where $Z=\int_{H} e^{-2 U(y)} \mu(d y)$ and $\mu$ is the Gaussian measure in $H$ with zero mean and covariance $Q=-\frac{1}{2} A^{-1}$. This

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is the reason to assume $A^{-1}$ of trace class. $Z$ is just a normalization constant in order to have a probability measure. Moreover system (1.2) is reversible; that is, if the law of $X(0)$ coincides with $v$, the reversed process $Y(t)=X(T-t), t \in[0, T]$ fulfills again (1.2); see, for example, [17]. In statistical mechanics $v$ is called a Gibbs measure.

The above assumptions do not guarantee well-posedness of problem (1.2)-(1.3); however, under suitable additional assumptions, a solution in a weak sense may be constructed, using the general strategy presented in [22] and applied in [12]. But in this paper we shall concentrate on the solutions of the Kolmogorov equation (1.1) only. The precise relation between the weak solution to (1.1) and the solution to (1.2)-(1.3) is established in the case of Lipschitz continuous $D U$, and in the example of Section 5. In such cases we prove that the expected formula

$$
u=\int_{0}^{+\infty} e^{-\lambda t} \mathbb{E}(f(X(t, \cdot))) d t
$$

holds for every $f \in C_{b}(H)$.
Throughout the paper we assume that $U$ belongs to a suitable Sobolev space. Then, the measure $v$ symmetrizes the operator

$$
\mathcal{K} u:=\frac{1}{2} \operatorname{Tr}\left[D^{2} u\right]+\langle A x-D U(x), D u\rangle,
$$

since for good functions $u, v$ (e.g., smooth cylindrical functions) we have

$$
\int_{H} \mathcal{K} u v d v=-\frac{1}{2} \int_{H}\langle D u, D v\rangle d v .
$$

Accordingly, we say that $u \in W^{1,2}(H, v)$ is a weak solution of equation (1.1) if

$$
\begin{equation*}
\lambda \int_{H} u \varphi d v+\frac{1}{2} \int_{H}\langle D u, D \varphi\rangle d v=\int_{H} f \varphi d \mu \quad \forall \varphi \in W^{1,2}(H, v) \tag{1.4}
\end{equation*}
$$

For every $\lambda>0$, the weak solutions to (1.1) when $f$ runs in $L^{2}(H, v)$ are precisely the elements of the domain of the self-adjoint realization $K$ of $\mathcal{K}$ associated to the quadratic form $(u, v) \mapsto \frac{1}{2} \int_{H}\langle D u, D \varphi\rangle d \nu$. See Section 3.1 for the definition of $K$.

Existence and uniqueness of a weak solution to (1.1) have been extensively studied, even in more general situations. We quote [1] for the Dirichlet form approach and [12] where it was proved that the restriction of $\mathcal{K}$ to exponential functions is essentially $m$-dissipative in $L^{2}(H, v)$. However, in all these papers only $W^{1,2}$ regularity of solutions was considered.

Our main concern is the investigation of the second derivative of the weak solution and of other maximal regularity results. In Section 3 we shall prove that the weak solution $u$ of equation (1.1) has the following properties:
(i) $u \in W^{2,2}(H, v)$,
(ii) $\int_{H}\left\|(-A)^{1 / 2} D u\right\|^{2} d v<\infty$,
and under further assumptions,

$$
\text { (iii) } \int_{H}\left\langle D^{2} U D u, D u\right\rangle d v<\infty
$$

Regularity of the second derivative of $u$ and sharp estimates for $D u$ are challenging problems for the theory of elliptic equations, even in finite dimensions. (i) Is a "natural" maximal regularity result for elliptic equations, both in finite and in infinite dimensions, while (ii) is typical of the infinite dimensional setting; see, for example, $[15,23]$ for the Ornstein-Uhlenbeck operator, when $U \equiv 0$. (iii) Is meaningful in the case that $D^{2} U$ is unbounded; otherwise it is contained in (i). It was known only in finite dimensions [19].

Properties (i)-(iii) allow us to study some perturbations of $\mathcal{K}$ of the type $\mathcal{K}_{1}=$ $\mathcal{K}+\mathcal{B}$, where

$$
\mathcal{B} u(x)=\langle B(x), D u(x)\rangle,
$$

and $B: H \rightarrow H$ is possibly unbounded. This is the subject of Section 4. Taking advantage of (i)-(iii), we can solve

$$
\begin{equation*}
\lambda u-K u-\langle B, D u\rangle=f \tag{1.5}
\end{equation*}
$$

under reasonable assumptions on $B$, when $\lambda$ is sufficiently large. The perturbed operator inherits some of the properties of $K$. For instance, it generates an analytic semigroup that preserves positivity. In some cases we can solve (1.5) for every $\lambda>0$, in a different $L^{2}$ setting. More precisely, adapting arguments from [14] that involve positivity preserving and compactness, we are able to prove the existence of $\rho \in L^{2}(H, v)$ such that a suitable realization of $\widetilde{K}_{1}$ of $\mathcal{K}_{1}$ is $m$-dissipative in $L^{2}(H, \zeta)$ where $\zeta(d x)=\rho(x) \nu(d x)$. Then, equation (1.5) can be solved for any $\lambda>0$ and any $f \in L^{2}(H, \zeta)$, and we prove that $\zeta$ is an invariant measure for the semigroup generated by $\widetilde{K}_{1}$ in $L^{2}(H, \zeta)$.

It is worth to note that $\mathcal{K}_{1}$ is the Kolmogorov operator corresponding to system

$$
\begin{equation*}
d X=(A X-D U(X)+B(X)) d t+d W(t), \quad X(0)=x \tag{1.6}
\end{equation*}
$$

which is not a gradient system in general. It may be useful in the study of nonequilibrium problems arising in statistical mechanics; see, for example, [18]. Another possible application of the regularity of the second derivative of the solution $u$ of (1.5) could be to the pathwise uniqueness of (1.6) (see the recent paper [11]), through the Veretennikov transform. This will be the object of future investigations.

In Sections 5 and 6 we show that the general theory may be applied to Kolmogorov equations of reaction-diffusion and Cahn-Hilliard stochastic PDEs.
2. Notation and preliminaries. In this section we fix notation and collect several preliminary results needed in the sequel. Though essentially known, they are scattered in different papers, so we will give details for the reader's convenience. Readers familiar with Sobolev spaces in infinite dimensions may jump to Section 3.

Let $H$ be a separable Hilbert space with inner product $\langle\cdot, \cdot\rangle$ and norm $\|\cdot\|$, endowed with a Gaussian measure $\mu:=\mathcal{N}_{0, Q}$ on the Borel sets of $H$, where $Q \in$ $\mathcal{L}(H)$ is a self-adjoint positive operator with finite trace. We choose once and for all an orthonormal basis $\left\{e_{k}: k \in \mathbb{N}\right\}$ of $H$ such that $Q e_{k}=\lambda_{k} e_{k}$ for $k \in \mathbb{N}$ and set $x_{k}=\left\langle x, e_{k}\right\rangle$ for each $x \in H$. We denote by $P_{n}$ the orthogonal projection on the linear span of $e_{1}, \ldots, e_{n}$. For each $k \in \mathbb{N} \cup\{+\infty\}$ we denote by $\mathcal{F} \mathcal{C}_{b}^{k}(H)$ the set of the cylindrical functions $\varphi(x)=\phi\left(x_{1}, \ldots, x_{n}\right)$ for some $n \in \mathbb{N}$, with $\phi \in C_{b}^{k}\left(\mathbb{R}^{n}\right)$.
2.1. Sobolev spaces with respect to $\mu$. For $p>1$ we set as usual $p^{\prime}:=p /(p-$ 1). If a function $\varphi: H \mapsto \mathbb{R}$ is Fréchet differentiable at $x \in H$, we denote by $D \varphi(x)$ its gradient at $x$. Moreover, we denote by $D_{k} \varphi(x)=\left\langle D \varphi(x), e_{k}\right\rangle$ its derivative in the direction of $e_{k}$, for every $k \in \mathbb{N}$.

For $0 \leq \theta \leq 1$ and $p>1$ the Sobolev spaces $W_{\theta}^{1, p}(H, \mu)$ are the completions of $\mathcal{F C} \mathcal{C}_{b}^{1}(H)$ in the Sobolev norms

$$
\|\varphi\|_{W_{\theta}^{1, p}(H, \mu)}^{p}:=\int_{H}\left(|\varphi|^{p}+\left\|Q^{\theta} D \varphi\right\|^{p}\right) d \mu=\int_{H}|\varphi|^{p}+\left(\sum_{k=1}^{\infty}\left(\lambda_{k}^{\theta} D_{k} \varphi\right)^{2}\right)^{p / 2} d \mu .
$$

For $\theta=1 / 2$ they coincide with the usual Sobolev spaces of the Malliavin Calculus; see, for example, [3], Chapter 5; for $\theta=0$ and $p=2$ they are the spaces considered in [15]. Such completions are identified with subspaces of $L^{p}(H, \mu)$ since the integration by parts formula

$$
\begin{align*}
& \int_{H} D_{k} \varphi \psi d \mu=-\int_{H} D_{k} \psi \varphi d \mu+\frac{1}{\lambda_{k}} \int_{H} x_{k} \varphi \psi d \mu,  \tag{2.1}\\
& \varphi, \psi \in \mathcal{F} \mathcal{C}_{b}^{1}(H),
\end{align*}
$$

allows us to easily show that the operators $Q^{\theta} D: \mathcal{F C}_{b}^{1}(H) \mapsto L^{p}(H, \mu ; H)$ are closable in $L^{p}(H, \mu)$, and the domains of their closures coincide with $W_{\theta}^{1, p}(H, \mu)$.

Moreover, since $x \mapsto x_{k} \in L^{s}(H, \mu)$ for every $s \geq 1$, (2.1) is extended by density to all $\varphi \in W_{\theta}^{1, q}(H, \mu), \psi \in W_{\theta}^{1, p}(H, \mu)$ such that $1 / p+1 / q<1$. In fact, extending [15], Lemma 9.2.7, to the case $p \geq 2$ it is possible to see that it holds for $1 / p+1 / q=1$ too.

The spaces $W_{\theta}^{1, p}(H, \mu ; H)$ are defined in a similar way, replacing $\mathcal{F \mathcal { C } _ { b } ^ { 1 } ( H ) \text { by }}$ linear combinations of functions of the type $\varphi e_{k}$, with $\varphi \in \mathcal{F} \mathcal{C}_{b}^{1}(H)$.
2.2. Sobolev spaces with respect to $v$. Concerning $U$ we shall assume the following:

HYPOTHESIS 2.1. $U: H \rightarrow \mathbb{R} \cup\{+\infty\}$ is convex, lower semicontinuous and bounded from below. Moreover $U \in W_{1 / 2}^{1,2}(H, \mu)$.

We denote by $v$ the log-concave measure $v(d x)=Z^{-1} e^{-2 U(x)} \mu(d x)$. Since $e^{-2 U}$ is bounded, $v(H)=1$.

Lemma 2.2. For every $p \geq 1, \mathcal{F C}_{b}^{\infty}(H)$ is dense in $L^{p}(H, v)$.
Proof. Since $H$ is separable, then $C_{b}(H)$ is dense in $L^{p}(H, v)$. Any $f \in$ $C_{b}(H)$ may be approached in $L^{p}(H, v)$ by the sequence $f_{n}(x):=f\left(P_{n} x\right)$, by the dominated convergence theorem. In its turn, the cylindrical functions $f_{n}$ are approached by their (finite dimensional) convolutions with smooth mollifiers that belong to $\mathcal{F C}_{b}^{\infty}(H)$.

We may apply the integration by parts formula (2.1) with $\psi$ replaced by $\psi e^{-2 U}$, that belongs to $W_{1 / 2}^{1,2}(H, \mu)$ for $\psi \in \mathcal{F} \mathcal{C}_{b}^{1}(H)$. We get, for $\varphi, \psi \in \mathcal{F} \mathcal{C}_{b}^{1}(H)$ and $h \in \mathbb{N}$,

$$
\begin{equation*}
\int_{H} D_{h} \varphi \psi d \nu+\int_{H} D_{h} \psi \varphi d \nu=2 \int_{H} D_{h} U \varphi \psi d \nu+\frac{1}{\lambda_{h}} \int_{H} x_{h} \varphi \psi d \nu \tag{2.2}
\end{equation*}
$$

Once again, the Sobolev spaces associated to the measure $v$ are introduced in a standard way with the help of the integration by parts formula (2.2). We recall that $\mathcal{L}_{2}(H)$ is the space of the Hilbert-Schmidt operators that are the bounded linear operators $L: H \mapsto H$ such that $\|L\|_{\mathcal{L}_{2}(H)}^{2}:=\sum_{h, k=1}^{\infty}\left\langle L e_{h}, e_{k}\right\rangle^{2}<\infty$.

Lemma 2.3. For all $q \geq 2$ the operators

$$
\begin{align*}
D: \mathcal{F} \mathcal{C}_{b}^{1}(H) & \mapsto L^{q}(H, v ; H) \\
Q^{ \pm 1 / 2} D: \mathcal{F} \mathcal{C}_{b}^{1}(H) & \mapsto L^{q}(H, v ; H)  \tag{2.3}\\
\left(D, D^{2}\right): \mathcal{F C}_{b}^{2}(H) & \mapsto L^{q}(H, v ; H) \times L^{q}\left(H, v ; \mathcal{L}_{2}(H)\right) \tag{2.4}
\end{align*}
$$

are closable in $L^{q}(H, v)$.
Proof. Let $\left(\varphi_{n}\right) \subset \mathcal{F} \mathcal{C}_{b}^{1}(H)$ converge to 0 in $L^{q}(H, v)$ and be such that $Q^{\theta} D \varphi_{n} \rightarrow W$ in $L^{q}(H, v ; H)$, with $\theta=0$ or $\theta=1 / 2$ or $\theta=-1 / 2$. Then for every $h \in \mathbb{N}$ the sequence $\left(\left\langle Q^{\theta} D \varphi_{n}, e_{h}\right\rangle\right)=\left(\lambda_{h}^{\theta} D_{h} \varphi_{n}\right)$ converges to $\left\langle W, e_{h}\right\rangle$ in $L^{q}(H, v)$. By formula (2.2) for each $\psi \in \mathcal{F} \mathcal{C}_{b}^{1}(H)$ we have

$$
\begin{equation*}
\int_{H} D_{h} \varphi_{n} \psi d v+\int_{H} D_{h} \psi \varphi_{n} d v=2 \int_{H} D_{h} U \varphi_{n} \psi d v+\frac{1}{\lambda_{k}} \int_{H} x_{h} \varphi_{n} \psi d v \tag{2.5}
\end{equation*}
$$

and letting $n \rightarrow \infty$, we get

$$
\lim _{n \rightarrow \infty} \int_{H} D_{h} \varphi_{n} \psi d v=\lim _{n \rightarrow \infty} \int_{H} \lambda_{h}^{-\theta}\left\langle W, e_{h}\right\rangle \psi d v=0
$$

Since $\mathcal{F} \mathcal{C}_{b}^{1}(H)$ is dense in $L^{q^{\prime}}(H, v)$, then $\left\langle W, e_{h}\right\rangle=0 v$-a.e. for every $h \in \mathbb{N}$, hence $W=0 v$-a.e., and the first statement is proved.

The proof of the second statement is similar. If $\left(\varphi_{n}\right) \subset \mathcal{F} \mathcal{C}_{b}^{2}(H)$ converge to 0 in $L^{q}(H, v)$ and $D \varphi_{n} \rightarrow W$ in $L^{q}(H, v ; H), D^{2} \varphi_{n} \rightarrow \mathcal{Q}$ in $L^{q}\left(H, v ; \mathcal{L}_{2}(H)\right)$, by the first part of the proof we have $W=0$, so that for every $k \in \mathbb{N}, D_{k} \varphi_{n} \rightarrow 0$ in $L^{q}(H, v)$. On the other hand, for each $h, k \in \mathbb{N},\left\langle D^{2} \varphi_{n} e_{h}, e_{k}\right\rangle=D_{h k} \varphi_{n}$ goes to $\left\langle\mathcal{Q} e_{h}, e_{k}\right\rangle$ in $L^{q}(H, v)$. Formula (2.2) applied to $D_{k} \varphi_{n}$ instead of $\varphi$ reads as
$\int_{H} D_{h k} \varphi_{n} \psi d v+\int_{H} D_{h} \psi D_{k} \varphi_{n} d v=2 \int_{H} D_{h} U D_{k} \varphi_{n} \psi d v+\frac{1}{\lambda_{k}} \int_{H} x_{k} D_{k} \varphi_{n} \psi d v$, for all $\psi \in \mathcal{F} \mathcal{C}_{b}^{1}(H)$. Letting $n \rightarrow \infty$ we get

$$
\lim _{n \rightarrow \infty} \int_{H} D_{h k} \varphi_{n} \psi d v=\lim _{n \rightarrow \infty} \int_{H}\left\langle\mathcal{Q} e_{h}, e_{k}\right\rangle \psi d v=0
$$

Then, $\left\langle\mathcal{Q} e_{h}, e_{k}\right\rangle=0$ a.e. for each $h$ and $k$, so that $\mathcal{Q}=0$, $v$-a.e.

REMARK 2.4. We remark that the restriction $q \geq 2$ comes from the integral $\int_{H} D_{h} U \varphi_{n} \psi d v$ in (2.5), where $D_{h} U \in L^{2}(H, v)$ as a consequence of Hypothesis 2.1. If $\|D U\| \in L^{p}(H, \mu)$ for some $p>2$ the proof of Lemma 2.3 works for any $q \geq p^{\prime}$.

DEFINITION 2.5. For $q \geq 2$ we still denote by $D, Q^{1 / 2} D, Q^{-1 / 2} D$, and by ( $D, D^{2}$ ) the closures in $L^{q}(H, v)$ of the operators defined in (2.3), (2.4).

We denote by $W^{1, q}(H, v)$ and by $W_{1 / 2}^{1, q}(H, v), W_{-1 / 2}^{1, q}(H, v)$, the domains of $D, Q^{1 / 2} D, Q^{-1 / 2} D$ in $L^{q}(H, v)$, respectively, and by $W^{2, q}(H, v)$ the domain of $\left(D, D^{2}\right)$ in $L^{q}(H, v)$.

Then, $W^{1, q}(H, v), W_{ \pm 1 / 2}^{1, q}(H, v)$ and $W^{2, q}(H, v)$ are Banach spaces with the norms

$$
\begin{aligned}
\|u\|_{W^{1, q}(H, v)}^{q} & =\int_{H}|u|^{q} d v+\int_{H}\|D u\|^{q} d \nu \\
\|u\|_{W_{ \pm 1 / 2}^{1, q}(H, v)}^{q} & =\int_{H}|u|^{q} d v+\int_{H}\left\|Q^{ \pm 1 / 2} D u\right\|^{q} d v \\
\|u\|_{W^{2, q}(H, v)}^{q} & =\|u\|_{W^{1, q}(H, v)}^{q}+\int_{H}\left\|D^{2} u\right\|_{\mathcal{L}_{2}(H)}^{q} d \nu
\end{aligned}
$$

Denoting by $D_{k} u:=\lambda_{k}^{-\theta}\left\langle Q^{\theta} D u, e_{k}\right\rangle$, with $\theta \in\{0,1 / 2,-1 / 2\}, D_{h k} u:=$ $\left\langle D^{2} u e_{h}, e_{k}\right\rangle$, the above Sobolev norms may be written in a more explicit way as

$$
\begin{aligned}
\|u\|_{W^{1, q}(H, v)}^{q} & =\int_{H}|u|^{q} d v+\int_{H}\left(\sum_{k \in \mathbb{N}}\left(D_{k} u\right)^{2}\right)^{q / 2} d v \\
\|u\|_{W_{ \pm 1 / 2}^{1, q}(H, v)}^{q} & =\int_{H}|u|^{q} d v+\int_{H}\left(\sum_{k \in \mathbb{N}} \lambda_{k}^{ \pm 1}\left(D_{k} u\right)^{2}\right)^{q / 2} d v, \\
\|u\|_{W^{2, q}(H, v)}^{q} & =\|u\|_{W^{1, q}(H, v)}^{q}+\int_{H}\left(\sum_{h, k \in \mathbb{N}}\left(D_{h k} u\right)^{2}\right)^{q / 2} d v \\
& =\|u\|_{W^{1, q}(H, v)}^{q}+\int_{H} \operatorname{Tr}\left(\left[D^{2} u\right]^{2}\right) d v
\end{aligned}
$$

For $q=2$, such spaces are Hilbert spaces with the respective scalar products

$$
\begin{aligned}
\langle u, v\rangle_{W^{1,2}(H, v)} & =\int_{H} u v d v+\int_{H} \sum_{k \in \mathbb{N}} D_{k} u D_{k} v d v, \\
\langle u, v\rangle_{W_{ \pm 1 / 2}^{1,2}(H, v)} & =\int_{H} u v d v+\int_{H} \sum_{k \in \mathbb{N}} \lambda_{k}^{ \pm 1} D_{k} u D_{k} v d v, \\
\langle u, v\rangle_{W^{2,2}(H, v)} & =\langle u, v\rangle_{W^{1,2}(H, v)}+\int_{H} \sum_{h, k \in \mathbb{N}} D_{h k} u D_{h k} v d v .
\end{aligned}
$$

REMARK 2.6. Let us make some remarks about the above definitions.
(1) It follows immediately from the definition that for every $u \in W^{1, p}(H, v)$ and $\varphi \in C_{b}^{1}(\mathbb{R})$, the superposition $\varphi \circ u$ belongs to $W^{1, p}(H, \nu)$, and $D(\varphi \circ u)=$ $\left(\varphi^{\prime} \circ u\right) D u$. This fact will be used frequently in the sequel.
(2) Formula (2.2) holds for each $\varphi \in \mathcal{F C}_{b}^{1}(H), \psi \in W^{1, q}(H, v)$ with $q \geq 2$. Indeed, it is sufficient to approach $\psi$ by a sequence of cylindrical functions in $\mathcal{F} \mathcal{C}_{b}^{1}(H)$, and to use (2.2) for the approximating functions, recalling that $D_{h} U$, $x_{h} \in L^{2}(H, v)$.
(3) Similarly, (2.2) holds for $\varphi \in W^{1, p}(H, v), \psi \in W^{1, q}(H, v)$ such that $1 / p+$ $1 / q \leq 1 / 2$.
2.2.1. Positive and negative parts of elements of $W^{1,2}(H, v)$. The following technical lemma will be used later to study positivity of solutions of (1.1).

Lemma 2.7. Let $u \in W^{1,2}(H, v)$. Then $|u|$ (and consequently, $u^{+}=$ $\left.\sup \{u, 0\}, u^{-}=\sup \{-u, 0\}\right)$ belongs to $W^{1,2}(H, v)$, and $D|u|=\operatorname{sign} u D u$. Moreover $D u=0$ a.e. in the set $u^{-1}(0)$, and $D u^{+}=D u \mathbb{1}_{\{u \geq 0\}}=D u \mathbb{1}_{\{u>0\}}, D u^{-}=$ $-D u \mathbb{1}_{\{u \leq 0\}}=-D u \mathbb{1}_{\{u<0\}}$.

Proof. Set $f_{n}(\xi)=\sqrt{\xi^{2}+1 / n}, \xi \in \mathbb{R}$. If $\left(u_{n}\right)$ is a sequence of functions in $\mathcal{F} \mathcal{C}_{b}^{1}(H)$ that approach $u$ in $W^{1,2}(H, v)$ and pointwise a.e., the functions $f_{n} \circ u_{n}$ belong to $\mathcal{F} \mathcal{C}_{b}^{1}(H)$ and approach $|u|$ in $W^{1,2}(H, v)$. Indeed, they converge to $|u|$ in $L^{2}(H, v)$ by the dominated convergence theorem, and $D\left(f_{n} \circ u_{n}\right)=f_{n}^{\prime} \circ u_{n} D u_{n}$ converge to $\operatorname{sign} u D u$ in $L^{2}(H, v ; H)$. The first statement follows.

Let us prove that $D u$ vanishes a.e. in the kernel of $u$. It is sufficient to prove that for every $u \in W^{1,2}(H, v)$ and $i \in \mathbb{N}$ we have

$$
\begin{equation*}
\int_{\{u=0\}} D_{i} u \varphi d v=0, \quad \varphi \in \mathcal{F} \mathcal{C}_{b}^{1}(H) \tag{2.6}
\end{equation*}
$$

Indeed, since $\mathcal{F} \mathcal{C}_{b}^{1}(H)$ is dense in $L^{2}(H, v)$, (2.6) implies that $D_{i} u \mathbb{1}_{\{u=0\}}$ is orthogonal to all elements of $L^{2}(H, v)$, hence it vanishes a.e.

Let $\theta: \mathbb{R} \mapsto \mathbb{R}$ be a smooth function with support contained in $[-1,1]$, with values in $[0,1]$ and such that $\theta(0)=1$. For $\varepsilon>0$ set $\theta_{\varepsilon}(\xi)=\theta(\xi / \varepsilon)$. The functions $\theta_{\varepsilon} \circ u$ have values in $[0,1]$ and converge pointwise to $\mathbb{1}_{\{u=0\}}$. Moreover, they belong to $W^{1,2}(H, v)$ and we have $D_{i}\left(\theta_{\varepsilon} \circ u\right)=\left(\theta_{\varepsilon}^{\prime} \circ u\right) D_{i} u=\left(\theta^{\prime} \circ u / \varepsilon\right) D_{i} u / \varepsilon$. Integrating we obtain

$$
\begin{aligned}
\int_{H} D_{i} u \varphi\left(\theta_{\varepsilon} \circ u\right) d v= & -\int_{H} u D_{i} \varphi\left(\theta_{\varepsilon} \circ u\right) d v \\
& -\int_{H} u \varphi D_{i}\left(\theta_{\varepsilon} \circ u\right) d v+2 \int_{H} u \varphi\left(\theta_{\varepsilon} \circ u\right) D_{i} U d v \\
& +\frac{1}{\lambda_{i}} \int_{H} x_{i} u \varphi\left(\theta_{\varepsilon} \circ u\right) d v
\end{aligned}
$$

As $\varepsilon \rightarrow 0$ we obtain by the dominated convergence theorem

$$
\begin{aligned}
\lim _{\varepsilon \rightarrow 0} \int_{H} D_{i} u \varphi\left(\theta_{\varepsilon} \circ u\right) d v & =\int_{\{u=0\}} D_{i} u \varphi d v, \\
\lim _{\varepsilon \rightarrow 0} \int_{H} u D_{i} \varphi\left(\theta_{\varepsilon} \circ u\right) d v & =\int_{\{u=0\}} u D_{i} \varphi d v=0, \\
\lim _{\varepsilon \rightarrow 0} \int_{H} u \varphi\left(\theta_{\varepsilon} \circ u\right) D_{i} U d v & =\int_{\{u=0\}} u \varphi D_{i} U d v=0, \\
\lim _{\varepsilon \rightarrow 0} \frac{1}{\lambda_{i}} \int_{H} x_{i} u \varphi\left(\theta_{\varepsilon} \circ u\right) d v & =\frac{1}{\lambda_{i}} \int_{\{u=0\}} x_{i} u \varphi d v=0 .
\end{aligned}
$$

The integral $\int_{H} u \varphi D_{i}\left(\theta_{\varepsilon} \circ u\right) d v$ vanishes too as $\varepsilon \rightarrow 0$, by the dominated convergence theorem. Indeed the support of $u \varphi D_{i}\left(\theta_{\varepsilon} \circ u\right)$ is contained in $u^{-1}([-\varepsilon, \varepsilon])$ so that its modulus is bounded by $\left\|\theta^{\prime}\right\|_{\infty}\|\varphi\|_{\infty}$. Moreover it converges to 0 pointwise as $\varepsilon \rightarrow 0$. So, letting $\varepsilon \rightarrow 0$ we obtain (2.6).

Once we know that $D u$ vanishes a.e. in the kernel of $u$, the formulas for $D u^{+}$ and $D u^{-}$follow from the equalities $u^{+}=(|u|+u) / 2, u^{-}=(|u|-u) / 2$.
2.2.2. Functional inequalities and embeddings. Under some additional assumptions important functional inequalities hold in the space $W^{1,2}(H, v)$.

Hypothesis 2.8. $U \in W_{0}^{1,2}(H, \mu)$ and $\|D U\| \in L^{p}(H, \mu)$ for some $p>2$.
We recall that since $A$ is invertible, and $-A^{-1}$ is nonnegative and compact, then

$$
-\omega:=\sup \{\langle A x, x\rangle: x \in D(A)\}<0
$$

Proposition 2.9. Let Hypotheses 2.1 and 2.8 hold. Then the following Poincaré and Logarithmic Sobolev inequalities hold:

$$
\begin{align*}
& \int_{H}\left(\varphi-\int_{H} \varphi d \nu\right)^{2} d \nu \leq \frac{1}{2 \omega} \int_{H}\|D \varphi\|^{2} d \nu, \quad \varphi \in W^{1,2}(H, \nu)  \tag{2.7}\\
& \int_{H} \varphi^{2} \log \left(\varphi^{2}\right) d v \leq \frac{1}{\omega} \int_{H}\|D \varphi\|^{2} d v+\int_{H} \varphi^{2} d \nu \log \left(\int_{H} \varphi^{2} d \nu\right)  \tag{2.8}\\
& \varphi \in W^{1,2}(H, v)
\end{align*}
$$

For the proof we refer to [15], Section 12.3.1.
Another useful property is the compact embedding of $W^{1,2}(H, v)$ in $L^{2}(H, v)$; see [10].

Proposition 2.10. Under Hypotheses 2.1 and $2.8, W^{1,2}(H, v)$ is compactly embedded in $L^{2}(H, v)$.

Proof. Let $\left(f_{n}\right)$ be a bounded sequence in $W^{1,2}(H, v)$. We look for a subsequence that converges in $L^{2}(H, v)$. By the Log-Sobolev inequality (2.8) the sequence is uniformly integrable, and hence it is sufficient to find a subsequence that converges almost everywhere.

The sequence $\left(f_{n} e^{-U}\right)$ is bounded in $W_{0}^{1, q}(H, \mu)$, with $q=2 p /(2+p) \in$ $(1,2)$. Indeed, it is bounded in $L^{2}(H, \mu)$, and hence it is bounded in $L^{q}(H, \mu)$, moreover $D\left(f_{n} e^{-U}\right)=D f_{n} e^{-U}-f_{n} D U e^{-U}$. Once again, $\left\|D f_{n} e^{-U}\right\|$ is bounded in $L^{2}(H, \mu)$, while the second addendum $f_{n} D U e^{-U}$ satisfies

$$
\begin{aligned}
\int_{H}\left\|f_{n} D U e^{-U}\right\|^{q} d \mu & \leq\left(\int_{H} f_{n}^{2} e^{-2 U} d \mu\right)^{q / 2}\left(\int_{H}\|D U\|^{2 q /(2-q)} d \mu\right)^{(2-q) / q} \\
& =\left\|f_{n}\right\|_{L^{2}(H, v)}^{q}\left(\int_{H}\|D U\|^{p} d \mu\right)^{(2-q) / q}
\end{aligned}
$$

so that it is bounded in $L^{q}(H, \mu)$.
Since the embedding $W_{0}^{1, q}(H, \mu) \subset L^{q}(H, \mu)$ is compact [5], there exists a subsequence that converges in $L^{q}(H, \mu)$ and a further subsequence that converges pointwise $\mu$-a.e. and also $v$-a.e., since $\nu$ is absolutely continuous with respect to $\mu$.
2.3. Moreau-Yosida approximations. An important tool in our analysis are the Moreau-Yosida approximations of $U$ defined for $\alpha>0$ by

$$
\begin{equation*}
U_{\alpha}(x)=\inf \left\{U(y)+\frac{|x-y|^{2}}{2 \alpha}, y \in H\right\}, \quad x \in H \tag{2.9}
\end{equation*}
$$

We recall that $U_{\alpha}(x) \leq U(x)$ and $U_{\alpha}(x)$ converges monotonically to $U(x)$ for each $x$ as $\alpha \rightarrow 0$. Moreover, each $U_{\alpha}$ is differentiable at any point, $D U_{\alpha}$ is Lipschitz continuous and $\left\|D U_{\alpha}\right\|$ converges monotonically to $\left\|D_{0} U\right\|$, at any $x$ such that the subdifferential of $U(x)$ is not empty. Here, $D_{0} U(x)$ is the element with minimal norm in the subdifferential of $U(x)$. At such points we have

$$
\begin{equation*}
\left\|D U_{\alpha}(x)-D_{0} U(x)\right\|^{2} \leq\left\|D_{0} U(x)\right\|^{2}-\left\|D U_{\alpha}(x)\right\|^{2} \tag{2.10}
\end{equation*}
$$

see, for example, [4], Chapter 2. If in addition $U \in C^{2}$, then $D_{0} U=D U$, and we have convergence of the second order derivatives, as the next lemma shows.

Lemma 2.11. Let $U: H \mapsto \mathbb{R}$ be convex and $C^{2}$. Then $\lim _{\alpha \rightarrow 0} D^{2} U_{\alpha}(x)=$ $D^{2} U(x)$ in $\mathcal{L}(H)$ for all $x \in H$.

Proof. For each $x \in H$ set $y_{\alpha}(x)=(I+\alpha D U)^{-1}(x)$, so that

$$
\begin{equation*}
y_{\alpha}(x)+\alpha D U\left(y_{\alpha}(x)\right)=x \tag{2.11}
\end{equation*}
$$

and by [4], Chapter 2,

$$
\begin{equation*}
D U_{\alpha}(x)=D U\left(y_{\alpha}\right) \tag{2.12}
\end{equation*}
$$

Since $U$ is convex, then $\left\langle D U(x)-D U\left(y_{\alpha}(x)\right), \alpha D U\left(y_{\alpha}(x)\right)\right\rangle=\langle D U(x)-$ $\left.D U\left(y_{\alpha}(x)\right), x-y_{\alpha}(x)\right\rangle \geq 0$. Taking the scalar product with $D U\left(y_{\alpha}(x)\right)$ yields $\left\|D U\left(y_{\alpha}(x)\right)\right\| \leq\|D U(x)\| /(1-\alpha)$, and letting $\alpha \rightarrow 0$ in (2.11) we get

$$
\lim _{\alpha \rightarrow 0} y_{\alpha}(x)=x \quad \forall x \in H
$$

Now it is clear that $y_{\alpha}$ is of class $C^{1}$, and differentiating (2.11) yields

$$
\begin{equation*}
y_{\alpha}^{\prime}(x)+\alpha D^{2} U\left(y_{\alpha}(x)\right) y_{\alpha}^{\prime}(x)=I . \tag{2.13}
\end{equation*}
$$

Since $U$ is convex,

$$
\left\|y_{\alpha}^{\prime}(x)\right\|_{\mathcal{L}(H)} \leq 1
$$

so that, letting $\alpha \rightarrow 0$ in (2.13) and recalling that $D^{2} U$ is continuous, we obtain

$$
\lim _{\alpha \rightarrow 0} y_{\alpha}^{\prime}(x)=I
$$

On the other hand, differentiating identity (2.12) gives $D^{2} U_{\alpha}(x)=D^{2} U\left(y_{\alpha}(x)\right)$. $y_{\alpha}^{\prime}(x)$ which yields the statement.
3. Elliptic problems. This section is devoted to the main result of the paper. In Section 3.1 we prove existence and uniqueness of a weak solution $u$ of equation (1.1). Section 3.2 is devoted to the particular case that $D U$ is Lipschitz continuous. This is an intermediate step in order to prove in Section 3.3 that under Hypothesis 2.1 we have

$$
u \in W^{2,2}(H, v) \cap W_{-1 / 2}^{1,2}(H, v)
$$

In Section 3.4 we show that if in addition $U$ is twice continuously differentiable, then

$$
\int_{H}\left\langle D^{2} U(x) D u(x), D u(x)\right\rangle v(d x)<\infty
$$

3.1. Weak solutions. We consider a Kolmogorov operator defined on $\mathcal{F C}_{b}^{2}(H)$ by

$$
\begin{equation*}
\mathcal{K} \varphi=\frac{1}{2} \operatorname{Tr}\left[D^{2} \varphi\right]+\frac{1}{2}\left\langle x, Q^{-1} D \varphi\right\rangle-\langle D U(x), D \varphi\rangle \tag{3.1}
\end{equation*}
$$

Using the partial derivatives $D_{k}$ and $D_{k k}, \mathcal{K}$ may be rewritten as

$$
\mathcal{K} \varphi(x)=\frac{1}{2} \sum_{k=1}^{\infty} D_{k k} \varphi(x)-\frac{1}{2} \sum_{k=1}^{\infty} \lambda_{k}^{-1} x_{k} D_{k} \varphi(x)-\sum_{k=1}^{\infty} D_{k} U(x) D_{k} \varphi(x)
$$

The measure $v$ enjoys the following important symmetrizing property:
Proposition 3.1. For all $\varphi \in \mathcal{F C}_{b}^{2}(H), \psi \in \mathcal{F} \mathcal{C}_{b}^{1}(H)$ we have

$$
\begin{equation*}
\int_{H} \mathcal{K} \varphi \psi d \nu=-\frac{1}{2} \int_{H}\langle D \varphi, D \psi\rangle d \nu \tag{3.2}
\end{equation*}
$$

Proof. Recalling (2.2) we get

$$
\begin{aligned}
\frac{1}{2} \int_{H} \sum_{k=1}^{\infty} D_{k k} \varphi(x) \psi(x) d v= & -\frac{1}{2} \int_{H} \sum_{k=1}^{\infty} D_{k} \varphi(x) D_{k} \psi(x) d v \\
& +\int_{H} \sum_{k=1}^{\infty}\left(D_{k} U(x) D_{k} \varphi(x)+\frac{1}{2 \lambda_{k}} x_{k} D_{k} \varphi(x)\right) d v
\end{aligned}
$$

and the conclusion follows (note that all series are finite sums in our case).
Let $f \in L^{2}(H, v), \lambda>0$. Taking into account formula (3.2), we say that $u \in$ $W^{1,2}(H, v)$ is a weak solution of equation (1.1) if we have

$$
\begin{equation*}
\lambda \int_{H} u \varphi d v+\frac{1}{2} \int_{H}\langle D u, D \varphi\rangle d v=\int_{H} f \varphi d v \quad \forall \varphi \in W^{1,2}(H, v) \tag{3.3}
\end{equation*}
$$

Since $\mathcal{F C}_{b}^{1}(H)$ is dense in $W^{1,2}(H, v)$, it is enough that the above equality is satisfied for every $\varphi \in \mathcal{F} \mathcal{C}_{b}^{1}(H)$.

The function $\mathcal{A}:\left(W^{1,2}(H, v)\right)^{2} \mapsto \mathbb{R}, \mathcal{A}(u, \varphi)=\lambda \int_{H} u \varphi d v+\frac{1}{2} \int_{H}\langle D u$, $D \varphi\rangle d \nu$ is bilinear, continuous and coercive, while the function $F: W^{1,2}(H, \nu) \mapsto$ $\mathbb{R}, F(\varphi)=\int_{H} f \varphi d \nu$, is linear and continuous. By the Lax-Milgram theorem there exists a unique $u \in W^{1,2}(H, v)$ such that $\mathcal{A}(u, \varphi)=F(\varphi)$ for each $\varphi \in$ $W^{1,2}(H, v)$; namely equation (1.1) has a unique weak solution $u \in W^{1,2}(H, v)$.

We denote by $K: D(K) \subset L^{2}(H, v) \mapsto L^{2}(H, v)$ the operator associated to the quadratic form $\mathcal{A}$ in $W^{1,2}(H, v)$. So, the domain $D(K)$ consists of all $u \in$ $W^{1,2}(H, v)$ such that there exists $v \in L^{2}(H, v)$ satisfying

$$
\frac{1}{2} \int_{H}\langle D u, D \varphi\rangle d v=-\langle v, \varphi\rangle_{L^{2}(H, v)}
$$

for all $\varphi \in W^{1,2}(H, v)$, or equivalently for all $\varphi \in \mathcal{F C}_{b}^{1}(H)$. In this case, $v=K u$. The weak solution $u$ to (1.1) belongs to $D(K)$, and it is just $(\lambda I-K)^{-1} f$.

REMARK 3.2. We have $\mathcal{F} \mathcal{C}_{b}^{2}(H) \subset D(K)$. In fact, for $u \in \mathcal{F} \mathcal{C}_{b}^{2}(H)$, integrating by parts we obtain

$$
\begin{equation*}
\frac{1}{2} \int_{H}\langle D u, D \varphi\rangle d v=-\int_{H}(\mathcal{K} u(x)) \varphi(x) v(d x) \tag{3.4}
\end{equation*}
$$

for all $\varphi \in \mathcal{F} \mathcal{C}_{b}^{1}(H)$. Here $\mathcal{K} u \in L^{2}(H, v)$ since it consists of the sum of a finite number of addenda, each of them in $L^{2}(H, v)$. Hence, $u \in D(K)$ and $K u=\mathcal{K} u$.

To study the domain of $K$ it is convenient to introduce a family of approximating problems, with $U$ replaced by its Moreau-Yosida approximations $U_{\alpha}$ defined in (2.9). Since $D U_{\alpha}$ is Lipschitz continuous, in the next section we consider the case of functions $U$ with Lipschitz gradient.
3.2. The case of Lipschitz continuous $D U$. Here we assume that $U: H \mapsto \mathbb{R}$ is a differentiable convex function bounded from below and with Lipschitz continuous gradient. Since $D U$ is Lipschitz, it has at most linear growth, and $U$ has at most quadratic growth. Therefore, it satisfies Hypothesis 2.1.

The aim of this section is to show that for every $f \in L^{2}(H, v)$ the weak solution to (1.1) belongs to $W^{2,2}(H, v) \cap W_{-1 / 2}^{1,2}(H, v)$ and the estimate

$$
\begin{align*}
& \lambda \int_{H}|D u|^{2} d v+\frac{1}{2} \int_{H} \operatorname{Tr}\left[\left(D^{2} u\right)^{2}\right] d v+\int_{H}\left\|Q^{-1 / 2} D u\right\|^{2} d v \\
& \quad+\int_{H}\left\langle D^{2} U D u, D u\right\rangle d v \leq 4 \int_{H} f^{2} d v \tag{3.5}
\end{align*}
$$

holds.
Note that $U \notin W^{2,2}(H, \mu)$ in general. The term $\left\langle D^{2} U D u, D u\right\rangle$ in the last integral is meant as follows: since $H$ is separable, and $\mu$ is nondegenerate, by [21], Theorem 6, DU:H$\mapsto H$ is Gateaux differentiable $v$ almost everywhere. The

Gateaux second order derivatives $D_{h k} U$ are bounded by a constant independent of $h, k$, since $D U$ is Lipschitz continuous so that the Lipschitz constant of each $D_{k} U$ is bounded by a constant independent of $k$. Since $u \in W_{-1 / 2}^{1,2}(H, v)$ the double series $\sum_{h, k} D_{h k} U D_{h} u D_{k} u$ is well defined and belongs to $L^{1}(H, v)$. Indeed,

$$
\begin{aligned}
\left|\sum_{h, k=1}^{\infty} D_{h k} U D_{h} u D_{k} u\right| & \leq C\left(\sum_{k=1}^{\infty}\left|D_{k} u\right|\right)^{2}=C\left(\sum_{k=1}^{\infty} \lambda_{k}^{-1 / 2}\left|D_{k} u\right| \lambda_{k}^{1 / 2}\right)^{2} \\
& \leq C\left\|Q^{-1 / 2} D u\right\|^{2} \operatorname{Tr} Q
\end{aligned}
$$

Moreover, we shall show that the weak solution is also a strong solution in the Friedrichs sense.

DEFINITION 3.3. A function $u \in L^{2}(H, v)$ is called strong solution (in the Friedrichs sense) to (1.1) if there is a sequence $\left(u_{n}\right)$ of $\mathcal{F} \mathcal{C}_{b}^{2}(H)$ functions that converge to $u$ in $L^{2}(H, v)$ and such that $\lambda u_{n}-\mathcal{K} u_{n} \rightarrow f$ in $L^{2}(H, v)$.

In fact, we begin with the strong solution. The procedure is the following: we show that the operator $\mathcal{K}: \mathcal{F} \mathcal{C}_{b}^{3}(H) \mapsto L^{2}(H, v)$ is dissipative, so that it is closable. Then we show that $(\lambda-\mathcal{K})\left(\mathcal{F} \mathcal{C}_{b}^{3}(H)\right)$ is dense in $L^{2}(H, v)$ for every $\lambda>0$. This implies that the closure $\overline{\mathcal{K}}$ of $\mathcal{K}$ generates a contraction semigroup in $L^{2}(H, v)$, and $\mathcal{F C} \mathcal{C}_{b}^{3}(H)$ is a core, that is, it is dense in $D(\overline{\mathcal{K}})$ endowed with the graph norm. In particular, for every $f \in L^{2}(H, v)$ and $\lambda>0$, equation (1.1) has a unique solution $u \in D(\overline{\mathcal{K}})$, which is a strong solution by definition. Then we show that $D(\overline{\mathcal{K}}) \subset W^{2,2}(H, v)$ and that (3.5) holds. Eventually, we prove that the strong solution coincides with the weak solution.
3.2.1. $\mathcal{K}: \mathcal{F C}_{b}^{3}(H) \mapsto L^{2}(H, v)$ is dissipative. This is just a simple consequence of the integration formula (3.4), taking $u=\varphi \in \mathcal{F C}{ }_{b}^{3}(H)$.
3.2.2. $(\lambda I-\mathcal{K})\left(\mathcal{F C}_{b}^{3}(H)\right)$ is dense in $L^{2}(H, v)$. We shall approach every element $f \in \mathcal{F} \mathcal{C}_{b}^{\infty}(H)$ by functions $g$ of the type $g=\lambda v-\mathcal{K} v$, first with $v \in \mathcal{F} \mathcal{C}_{b}^{2}(H)$ and then with $v \in \mathcal{F C}_{b}^{3}(H)$. This will be done using existence and regularity results for differential equations in finite dimensions. Since $\mathcal{F} \mathcal{C}_{b}^{\infty}(H)$ is dense in $L^{2}(H, v)$, our aim will be achieved.

We recall that $P_{n}$ is the orthogonal projection on the linear span of $e_{1}, \ldots, e_{n}$. We identify $P_{n}(H)$ with $\mathbb{R}^{n}$, by the obvious isomorphism $\mathbb{R}^{n} \mapsto P_{n}(H), \xi \mapsto$ $\sum_{k=1}^{n} \xi_{k} e_{k}$. The induced Gaussian measure in $\mathbb{R}^{n}$ is just $\mathcal{N}_{0, Q_{n}}$ where $Q_{n}=$ $\operatorname{diag}\left(\lambda_{1}, \ldots, \lambda_{n}\right)$.

For any function $v: H \mapsto \mathbb{R}$ we identify $v \circ P_{n}$ with the function $v_{n}: \mathbb{R}^{n} \mapsto \mathbb{R}$, $v_{n}(\xi):=v\left(\sum_{k=1}^{n} \xi_{k} e_{k}\right)$. In particular, we identify $U \circ P_{n}: H \mapsto \mathbb{R}$ with the function $U_{n}: \mathbb{R}^{n} \mapsto \mathbb{R}, U_{n}(\xi):=U\left(\sum_{k=1}^{n} \xi_{k} e_{k}\right)$. $U_{n}$ is convex, and $D U_{n}$ is Lipschitz continuous, and hence $U_{n}$ belongs to $W^{2, \infty}\left(\mathbb{R}^{n}, d \xi\right) \subset W^{2, \infty}\left(\mathbb{R}^{n}, \mathcal{N}_{0}, Q_{n}\right)$.

For $\lambda>0$ let us consider the problem

$$
\begin{equation*}
\lambda v_{n}-\mathcal{L} v_{n}+\left\langle D U_{n}, D v_{n}\right\rangle=f_{n} \tag{3.6}
\end{equation*}
$$

where the Ornstein-Uhlenbeck operator $\mathcal{L}$ in $\mathbb{R}^{n}$ is defined by

$$
\mathcal{L} \varphi(\xi)=\frac{1}{2} \sum_{k=1}^{n}\left(D_{k k} \varphi(\xi)-\lambda_{k}^{-1} \xi_{k} D_{k} \varphi(\xi)\right), \quad \xi \in \mathbb{R}^{n}
$$

Since $D U_{n}$ is Lipschitz continuous, (3.6) has a unique solution $v_{n} \in \bigcup_{\alpha \in(0,1)}$ $C_{b}^{2+\alpha}\left(\mathbb{R}^{n}\right)$. A reference is [20], Theorem 1. In fact [20], Theorem 1, deals with large $\lambda$ 's, but a standard application of the maximum principle (e.g., [20], Lemma 2.4) and of the Schauder estimates of [20], Theorem 1, show that (3.6) is uniquely solvable in $C_{b}^{2+\theta}\left(\mathbb{R}^{n}\right)$ for each $\lambda>0$. Moreover, an estimate for the first order derivatives of $v_{n}$,

$$
\begin{equation*}
\left\|\left|D v_{n}\right|\right\|_{\infty} \leq \frac{1}{\lambda}\left\|\left|D f_{n}\right|\right\|_{\infty} \tag{3.7}
\end{equation*}
$$

follows from the well-known probabilistic representation formula for $v_{n}$,

$$
\begin{equation*}
v_{n}(\xi)=\int_{0}^{\infty} e^{-\lambda t} \mathbb{E}\left(f\left(X_{n}(t, \xi)\right)\right) d t, \quad \xi \in \mathbb{R}^{n} \tag{3.8}
\end{equation*}
$$

$X_{n}(t, \xi)$ being the solution to the stochastic ode in $\mathbb{R}^{n}$

$$
\left\{\begin{array}{l}
d X_{n}(t, \xi)=-\frac{1}{2} Q_{n}^{-1} X_{n}(t, \xi) d t-D U_{n}\left(X_{n}(t, \xi)\right) d t+d W_{n}(t) \\
X_{n}(0, \xi)=\xi
\end{array}\right.
$$

where $W_{n}(t)=P_{n} W(t)$ is a standard Brownian motion in $\mathbb{R}^{n}$. Indeed, (3.7) follows taking into account that

$$
\begin{aligned}
d\left(X_{n}(t, x)-X_{n}(t, y)\right)=-\frac{1}{2} & \left(Q_{n}^{-1}\left(X_{n}(t, x)-X_{n}(t, y)\right) d t\right. \\
& \left.-\left(D U_{n}\left(X_{n}(t, x)\right)-D U_{n}\left(X_{n}(t, y)\right)\right)\right) d t
\end{aligned}
$$

so that $X_{n}(\cdot, x)-X_{n}(\cdot, y)$ is almost surely differentiable, and taking the scalar product by $X_{n}(t, x)-X_{n}(t, y)$ we get $\frac{d}{d t}\left\|X_{n}(t, x)-X_{n}(t, y)\right\|^{2} \leq 0$, by the monotonicity of $D U_{n}$. This implies $\left\|X_{n}(t, x)-X_{n}(t, y)\right\| \leq\|x-y\|$ and consequently $\left|v_{n}^{\varepsilon}(x)-v_{n}^{\varepsilon}(y)\right| \leq\left\|f_{n}\right\|_{\text {Lip }}\|x-y\| / \lambda$.

Going back to infinite dimensions, we set

$$
\begin{equation*}
V_{n}(x):=v_{n}\left(x_{1}, \ldots, x_{n}\right), \quad x \in H . \tag{3.9}
\end{equation*}
$$

Then $V_{n} \in \mathcal{F} \mathcal{C}_{b}^{2}(H)$, and

$$
\begin{equation*}
\lambda V_{n}-\mathcal{K} V_{n}=f \circ P_{n}+\left\langle D U-D\left(U \circ P_{n}\right), D V_{n}\right\rangle, \tag{3.10}
\end{equation*}
$$

where $f \circ P_{n}=f$ for $n$ large enough, since $f$ is cylindrical. The right-hand side converges to $f$ as $n \rightarrow \infty$ since estimate (3.7) implies
$\left|\left\langle D U(x)-D\left(U \circ P_{n}\right)(x), D V_{n}^{\varepsilon}(x)\right\rangle\right| \leq \frac{1}{\lambda} \sup _{y \in H}\|D f(y)\|\left\|D U(x)-D\left(U \circ P_{n}\right)(x)\right\|$,
which goes to 0 pointwise, since $D U$ is continuous, and in $L^{2}(H, v)$ by the dominated convergence theorem, since

$$
\left\|D\left(U \circ P_{n}\right)(x)\right\| \leq[D U]_{\mathrm{Lip}}\left\|P_{n} x\right\|+\|D U(0)\| \leq[D U]_{\mathrm{Lip}}\|x\|+\|D U(0)\|,
$$

for each $n \in \mathbb{N}$. Therefore, $\lambda V_{n}-\mathcal{K} V_{n}$ converges to $f$ in $L^{2}(H, v)$, which implies that $(\lambda I-\mathcal{K})\left(\mathcal{F} \mathcal{C}_{b}^{2}(H)\right)$ is dense in $L^{2}(H, v)$.

This will be used later, in the proof of Proposition 3.8; however, it is not enough for our aims. This is because the formula, (3.20), which is the starting point of all our optimal estimates, is obtained differentiating $\lambda u-\mathcal{K} u$ for a cylindrical $u$, and we need that $u$ has third order derivatives. So, we shall approximate using $\mathcal{F C} \mathcal{C}_{b}^{3}$ functions instead of only $\mathcal{F C} \mathcal{C}_{b}^{2}$ functions.

To be able to use regularity theorems for elliptic equations in $\mathbb{R}^{n}$ that yield $C^{3}$ solutions, we need regular coefficients, so we approach $U_{n}$ in a standard way by convolution with smooth mollifiers. Precisely, we fix once and for all a function $\theta \in C_{c}^{\infty}\left(\mathbb{R}^{n}\right)$ with support contained in the ball $B(0,1)$ of center 0 and radius 1 , such that $\int_{\mathbb{R}^{n}} \theta(\xi) d \xi=1$, and for $\varepsilon>0$ we set

$$
U_{n}^{\varepsilon}(\xi)=\int_{\mathbb{R}^{n}} U_{n}(\xi-\varepsilon y) \theta(y) d y, \quad \xi \in \mathbb{R}^{n}
$$

Then $U_{n}^{\varepsilon}$ is smooth and convex, and $D U_{n}^{\varepsilon}$ is Lipschitz continuous. Moreover,

$$
\begin{align*}
\left|D U_{n}(\xi)-D \mathcal{U}_{n}^{\varepsilon}(\xi)\right| & =\left|\int_{\mathbb{R}^{n}}\left(D U_{n}(\xi)-D U_{n}(\xi-\varepsilon y)\right) \theta(y) d y\right| \\
& \leq \varepsilon\left[D U_{n}\right]_{\mathrm{Lip}} \int_{\mathbb{R}^{n}}|y| \theta(y) d y  \tag{3.11}\\
& \leq \varepsilon\left[D U_{n}\right]_{\mathrm{Lip}} \leq \varepsilon[D U]_{\mathrm{Lip}}, \quad \xi \in \mathbb{R}^{n}
\end{align*}
$$

For $\lambda>0$ and $\varepsilon>0$ let us consider the problem

$$
\begin{equation*}
\lambda v_{n}^{\varepsilon}-\mathcal{L} v_{n}^{\varepsilon}+\left\langle D U_{n}^{\varepsilon}, D v_{n}^{\varepsilon}\right\rangle=f_{n} \tag{3.12}
\end{equation*}
$$

As before, since $D U_{n}^{\varepsilon}$ are Lipschitz continuous, (3.12) has a unique solution $v_{n}^{\varepsilon} \in$ $\bigcup_{\alpha \in(0,1)} C_{b}^{2+\alpha}\left(\mathbb{R}^{n}\right)$, again by [20], Theorem 1 . The functions $v_{n}^{\varepsilon}$ are represented by

$$
\begin{equation*}
v_{n}^{\varepsilon}(x)=\int_{0}^{\infty} e^{-\lambda t} \mathbb{E}\left(f_{n}\left(X^{\varepsilon}(t, x)\right)\right) d t \tag{3.13}
\end{equation*}
$$

where $X^{\varepsilon}(t, x)$ is the solution to the stochastic ode

$$
\left\{\begin{array}{l}
d X^{\varepsilon}(t, x)=-\frac{1}{2} Q_{n}^{-1} X^{\varepsilon}(t, x) d t-D U_{n}^{\varepsilon}\left(X^{\varepsilon}(t, x)\right) d t+d W_{n}(t) \\
X^{\varepsilon}(0, x)=x
\end{array}\right.
$$

and $W_{n}(t)$ is a standard Brownian motion in $\mathbb{R}^{n}$. The representation formula (3.13) yields the sup norm estimates

$$
\begin{align*}
\left\|v_{n}^{\varepsilon}\right\|_{\infty} & \leq \frac{1}{\lambda}\left\|f_{n}\right\|_{\infty},  \tag{3.14}\\
\left\|\left|D v_{n}^{\varepsilon}\right|\right\|_{\infty} & \leq \frac{1}{\lambda}\left\|\left|D f_{n}\right|\right\|_{\infty} . \tag{3.15}
\end{align*}
$$

Equation (3.14) is immediate, while (3.15) follows arguing as in the proof of (3.7), since $D U_{n}^{\varepsilon}$ is monotonic as well.

We want to show that $v_{n}^{\varepsilon} \in C_{b}^{3}\left(\mathbb{R}^{n}\right)$. Since $D U_{n}^{\varepsilon}$ is smooth, then $v_{n}^{\varepsilon}$ belongs to $C^{\infty}\left(\mathbb{R}^{n}\right)$ by local elliptic regularity, and we need only to prove that its third order derivatives are bounded. To this end we differentiate both sides of (3.12) with respect to $x_{i}$, getting

$$
\lambda D_{i} v_{n}^{\varepsilon}-\mathcal{L} D_{i} v_{n}^{\varepsilon}+\frac{1}{\lambda_{i}} D_{i} v_{n}^{\varepsilon}+\left\langle D U_{n}^{\varepsilon}, D\left(D_{i} v_{n}^{\varepsilon}\right)\right\rangle=D_{i} f_{n}-\left\langle D\left(D_{i} U_{n}^{\varepsilon}\right), D v_{n}^{\varepsilon}\right\rangle
$$

The right-hand side is Hölder continuous and bounded. Applying once again the Schauder theorem [20], Theorem 1, we obtain $D_{i} v_{n}^{\varepsilon} \in C_{b}^{2+\alpha}\left(\mathbb{R}^{n}\right)$ for each $\alpha \in$ $(0,1)$. In particular, $v_{n}^{\varepsilon} \in C_{b}^{3}\left(\mathbb{R}^{n}\right)$.

Let us go back to infinite dimensions and set

$$
\begin{equation*}
V_{n}^{\varepsilon}(x):=v_{n}^{\varepsilon}\left(x_{1}, \ldots, x_{n}\right), \quad \mathcal{U}_{n}^{\varepsilon}(x)=U_{n}^{\varepsilon}\left(x_{1}, \ldots, x_{n}\right), \quad x \in H \tag{3.16}
\end{equation*}
$$

Then $V_{n}^{\varepsilon} \in \mathcal{F} \mathcal{C}_{b}^{3}(H)$ and

$$
\begin{equation*}
\lambda V_{n}^{\varepsilon}-\mathcal{K} V_{n}^{\varepsilon}=f \circ P_{n}+\left\langle D U-D \mathcal{U}_{n}^{\varepsilon}, D V_{n}\right\rangle . \tag{3.17}
\end{equation*}
$$

Concerning the right-hand side, taking into account (3.15) and (3.11), we get

$$
\begin{aligned}
& \left|\left\langle D U(x)-D \mathcal{U}_{n}^{\varepsilon}(x), D V_{n}^{\varepsilon}(x)\right\rangle\right| \\
& \quad \leq \frac{1}{\lambda} \sup _{y \in H}\|D f(y)\|\left(\left\|D U(x)-D\left(U \circ P_{n}\right)(x)\right\|\right. \\
& \left.\quad+\left\|D\left(U \circ P_{n}\right)(x)-D \mathcal{U}_{n}^{\varepsilon}(x)\right\|\right) \\
& \quad \leq \frac{1}{\lambda} \sup _{y \in H}\|D f(y)\|\left(\left\|D U(x)-D\left(U \circ P_{n}\right)(x)\right\|+\varepsilon[D U]_{\operatorname{Lip}(X)}\right)
\end{aligned}
$$

so that

$$
\begin{aligned}
& \left\|\left\langle D U-D \mathcal{U}_{n}^{\varepsilon}, D V_{n}^{\varepsilon}\right\rangle\right\|_{L^{2}(H, v)}^{2} \\
& \quad \leq\left(\frac{1}{\lambda} \sup _{y \in H}\|D f(y)\|\right)^{2} 2\left(\int_{H}\left\|D U-D\left(U \circ P_{n}\right)\right\|^{2} d v+\left(\varepsilon[D U]_{\operatorname{Lip}(X)}\right)^{2}\right)
\end{aligned}
$$

where the first integral $\int_{H}\left\|D U-D\left(U \circ P_{n}\right)\right\|^{2} d v$ vanishes as $n \rightarrow \infty$, as we already remarked. Therefore, $\left\|\left\langle D U-D \mathcal{U}_{n}^{\varepsilon}, D V_{n}^{\varepsilon}\right\rangle\right\|_{L^{2}(H, v)}$ is as small as we wish provided we take $n$ large and $\varepsilon$ small, and the same holds for $\lambda V_{n}^{\varepsilon}-\mathcal{K} V_{n}^{\varepsilon}-f$.

Summarizing, we have proved the following proposition.

Proposition 3.4. The closure $\overline{\mathcal{K}}$ of the operator $\mathcal{K}: \mathcal{F C}_{b}^{3}(H) \mapsto L^{2}(H, \nu)$ is $m$-dissipative, so that it generates a strongly continuous contraction semigroup in $L^{2}(H, v)$. In particular, for every $\lambda>0$ and $f \in L^{2}(H, v)$ problem (1.1) has a unique strong solution $u$, that is: there is a sequence $\left(u_{n}\right) \subset \mathcal{F} \mathcal{C}_{b}^{3}(H)$ such that $u_{n} \rightarrow u$ and $\lambda u_{n}-\mathcal{K} u_{n} \rightarrow f$ in $L^{2}(H, v)$.
3.2.3. $W^{2,2}(H, v)$ regularity of the strong solution and other estimates. To prove our estimates it is sufficient to consider functions $u \in \mathcal{F C} \mathcal{C}_{b}^{3}(H)$, which is dense in the domain of $\overline{\mathcal{K}}$. So, we fix $u \in \mathcal{F} \mathcal{C}_{b}^{3}(H), \lambda>0$, and we set

$$
\lambda u-\mathcal{K} u=f
$$

Estimates on $u$ and on $D u$ in terms of $f$ are elementary. They are obtained multiplying both sides by $u$ and taking into account (3.2).

Lemma 3.5. We have

$$
\int_{H}\left(\lambda u^{2}+\frac{1}{2}\|D u\|^{2}\right) d v=\int_{H} u f d v
$$

and therefore

$$
\begin{equation*}
\int_{H} u^{2} d v \leq \frac{1}{\lambda^{2}} \int_{H} f^{2} d v \tag{3.18}
\end{equation*}
$$

and

$$
\begin{equation*}
\int_{H}\|D u\|^{2} d v \leq \frac{2}{\lambda} \int_{H} f^{2} d \nu \tag{3.19}
\end{equation*}
$$

Estimates on the second order derivatives are less obvious. They are a consequence of the following proposition.

Proposition 3.6. For each $u \in \mathcal{F C}_{b}^{3}(H)$ we have

$$
\begin{align*}
& \lambda \int_{H}\|D u\|^{2} d v+\frac{1}{2} \int_{H} \operatorname{Tr}\left[\left(D^{2} u\right)^{2}\right] d v+\frac{1}{2} \int_{H}\left\|Q^{-1 / 2} D u\right\|^{2} d v \\
& \quad+\int_{H}\left\langle D^{2} U D u, D u\right\rangle d v=\int_{H}\langle D u, D f\rangle d v=2 \int_{H}(\lambda u-f) f d v \tag{3.20}
\end{align*}
$$

Proof. As in Section 3.2.2, we differentiate the equality $\lambda u-\mathcal{K} u=f$ with respect to $x_{i}$, then we multiply by $D_{i} u$ and sum up. We obtain

$$
\lambda\|D u\|^{2}-\sum_{i=1}^{\infty}\left(\mathcal{K} D_{i} u\right) D_{i} u+\sum_{i=1}^{\infty} \frac{\left(D_{i} u\right)^{2}}{2 \lambda_{i}}+\sum_{i, j=1}^{\infty} D_{i j} U D_{i} u D_{j} u=\langle D f, D u\rangle
$$

where the series are in fact finite sums. Integrating on $H$ and taking (3.1) into account, (3.20) follows.

As a corollary of Lemma 3.5 and Proposition 3.6 we obtain estimates on the strong solution to (1.1).

Proposition 3.7. Let $\lambda>0, f \in L^{2}(H, v)$, and let $u$ be the strong solution to (1.1). Then $u \in W^{2,2}(H, v) \cap W_{-1 / 2}^{1,2}(H, v)$, and

$$
\begin{align*}
& \lambda \int_{H}\|D u\|^{2} d v+\frac{1}{2} \int_{H} \operatorname{Tr}\left[\left(D^{2} u\right)^{2}\right] d v+\frac{1}{2} \int_{H}\left\|Q^{-1 / 2} D u\right\|^{2} d v \\
& \quad+\int_{H}\left\langle D^{2} U D u, D u\right\rangle d v \leq 4 \int_{H} f^{2} d v . \tag{3.21}
\end{align*}
$$

In addition, if $f \in \mathcal{F C}_{b}^{\infty}(H)$, then $u$ is $v$-essentially bounded, and we have

$$
\begin{equation*}
\text { ess } \sup _{x \in H}|u(x)| \leq \frac{1}{\lambda} \sup _{x \in H}|f(x)| \tag{3.22}
\end{equation*}
$$

Proof. Let $u_{j} \in \mathcal{F} \mathcal{C}_{b}^{3}(H)$ approach $u$ in $D(\overline{\mathcal{K}})$. By estimate (3.19), $D u_{j} \rightarrow$ $D u$ in $L^{2}(H, v ; H)$. By Proposition 3.6, equality (3.20) holds, with $u_{j}$ replacing $u$, and $f_{j}:=\lambda u_{j}-\mathcal{K} u_{j}$ replacing $f$. Then

$$
\begin{aligned}
& \lambda \int_{H}\left\|D u_{j}\right\|^{2} d v+\frac{1}{2} \int_{H} \operatorname{Tr}\left[\left(D^{2} u_{j}\right)^{2}\right] d v+\frac{1}{2} \int_{H}\left\|Q^{-1 / 2} D u_{j}\right\|^{2} d v \\
& \quad+\int_{H}\left\langle D^{2} U D u_{j}, D u_{j}\right\rangle d v \leq 2 \int_{H}\left(\lambda u_{j}-f_{j}\right) f_{j} d v \leq 4\left\|f_{j}\right\|_{L^{2}(H, v)}^{2}
\end{aligned}
$$

while by (3.18) we have $\lambda\left\|u_{j}\right\|_{L^{2}(H, v)} \leq\left\|f_{j}\right\|_{L^{2}(H, v)}$. Since $f_{j} \rightarrow f$ in $L^{2}(H, v)$ as $j \rightarrow \infty,\left(u_{j}\right)$ is a Cauchy sequence in $W^{2,2}(H, v)$ and in $W_{-1 / 2}^{1,2}(H, v)$. So, $u$ belongs to such spaces, and letting $j \rightarrow \infty$ estimate (3.21) follows.

To prove the last statement, for $f \in \mathcal{F C}_{b}^{\infty}(H)$ we approach $u$ by the functions used in the proof of Proposition 3.4. Then (3.22) follows from (3.14), taking into account that for a suitable sequence $\left(j_{k}\right),\left(u_{j_{k}}\right)$ converges to $u$, $v$-a.e.
3.2.4. Weak $=$ strong. For $\lambda>0$ and $f \in L^{2}(H, v)$, let $u$ be the strong solution to (1.1) given by Proposition 3.4. Let $u_{n} \in \mathcal{F C}_{b}^{3}(H)$ be such that $u_{n} \rightarrow u$ and $f_{n}:=\lambda u_{n}-\mathcal{K} u_{n} \rightarrow f$ in $L^{2}(H, \nu)$. As we remarked in the proof of Proposition 3.7, $u_{n} \rightarrow u$ in $W^{1,2}(H, v)$.

Fix $\varphi \in \mathcal{F} \mathcal{C}_{b}^{1}(H)$. Multiplying both sides of $\lambda u_{n}-\mathcal{K} u_{n}=f_{n}$ by $\varphi$, integrating over $H$ and recalling (3.2), we obtain

$$
\lambda \int_{H} u_{n} \varphi d v+\frac{1}{2} \int_{H}\left\langle D u_{n}, D \varphi\right\rangle d v=\int_{H} f_{n} \varphi d \nu
$$

Letting $n \rightarrow \infty$ yields that $u$ is the weak solution to (1.1). So, weak and strong solutions to (1.1) do coincide.

As a consequence of coincidence of strong and weak solutions we obtain a probabilistic representation formula for the weak solution to (1.1). Let $W(t)$ be any $H$-valued cylindrical Wiener process defined in a probability space $(\Omega, \mathcal{F}, \mathbb{P})$. A construction of such a process may be found, for example, in [13], Section 4.3. For each $x \in H$ consider the stochastic differential equation

$$
\begin{equation*}
d X=(A X-D U(X)) d t+d W(t), \quad X(0)=x \tag{3.23}
\end{equation*}
$$

We recall that a mild solution to (3.23) is a $\mathcal{F}_{t}$ adapted, $H$-continuous process that satisfies

$$
X(t)=e^{t A} x-\int_{0}^{t} e^{(t-s) A} D U(X(s)) d s+\int_{0}^{t} e^{(t-s) A} d W(s), \quad t \geq 0
$$

where $\mathcal{F}_{t}$ is the natural filtration of $W(t)$. Existence and uniqueness of a mild solution to (3.23) follow, for example, from [14], Theorem 5.5.8; see also Remark 5.5.7 of [14].

Proposition 3.8. For $\lambda>0$ and $f \in C_{b}(H)$, let $u$ be the weak solution to (1.1). Then

$$
\begin{equation*}
u=\int_{0}^{+\infty} e^{-\lambda t} \mathbb{E} f(X(t, \cdot)) d t \tag{3.24}
\end{equation*}
$$

Proof. As a first step, let $f \in \mathcal{F} C_{b}^{\infty}(H)$, let $V_{n}$ be the functions defined in (3.9) and set $f_{n}:=\lambda V_{n}-K V_{n}$. In Section 3.2.2 we have shown that $\lim _{n \rightarrow \infty} f_{n}=f$ in $L^{2}(H, v)$. Therefore, $u=R(\lambda, K) f=\lim _{n \rightarrow \infty} R(\lambda, K) f_{n}=$ $\lim _{n \rightarrow \infty} V_{n}$. On the other hand, we have $V_{n}(x)=v_{n}\left(x_{1}, \ldots, x_{n}\right)$, where the functions $v_{n}$ solve (3.6). This implies that $V_{n}$ satisfies

$$
\begin{equation*}
V_{n}(x)=\int_{0}^{+\infty} e^{-\lambda t} \mathbb{E} f\left(X_{n}(t, x)\right) d t, \quad x \in H \tag{3.25}
\end{equation*}
$$

where $X_{n}$ is the mild solution to

$$
\begin{equation*}
d X_{n}=\left(A X_{n}-D\left(U \circ P_{n}\right)\left(X_{n}\right)\right) d t+d W(t), \quad X_{n}(0)=P_{n} x \tag{3.26}
\end{equation*}
$$

and for every $t>0, x \in X$ we have $\lim _{n \rightarrow \infty} X_{n}(t, x)=X(t, x)$, a.s. Letting $n \rightarrow \infty$ in (3.25), the left-hand side goes to $u$ in $L^{2}(H, v)$. The right-hand side converges to $\int_{0}^{+\infty} e^{-\lambda t} \mathbb{E} f(X(t, x)) d t$ pointwise and in $L^{2}(H, v)$ by the dominated convergence theorem. Indeed, for each $x \in H$ and $t>0$ we have $\lim _{n \rightarrow \infty} f\left(X_{n}(t, x)\right)=f(X(t, x))$ a.s., and $\left|f\left(X_{n}(t, x)\right)\right| \leq\|f\|_{\infty}$. Therefore, the statement holds if $f \in \mathcal{F} C_{b}^{\infty}(H)$.

If $f \in C_{b}(H)$, it is possible to approach it, pointwise and in $L^{2}(H, v)$, by a sequence $\left(f_{n}\right)$ of functions belonging to $\mathcal{F} C_{b}^{\infty}(H)$. For instance, one can take approximations by convolution of $f \circ P_{n}$. Then, $u_{n}:=R(\lambda, K) f_{n}$ satisfy (3.24) with $f$ replaced by $f_{n}$ and converge to $u=R(\lambda, K) f$ in $L^{2}(H, v)$. The right-hand sides converge to $\int_{0}^{+\infty} e^{-\lambda t} \mathbb{E} f(X(t, \cdot)) d t$ in $L^{2}(H, \nu)$, again by the dominated convergence theorem, and the statement follows.
3.3. The general case. Here we apply the results of Section 3.2 to prove our main result.

THEOREM 3.9. Under Hypothesis 2.1, for every $\lambda>0$ and $f \in L^{2}(H, v)$, the weak solution $u$ to (1.1) belongs to $W^{2,2}(H, v) \cap W_{-1 / 2}^{1,2}(H, v)$, and it satisfies

$$
\begin{align*}
\int_{H} u^{2} d v & \leq \frac{1}{\lambda^{2}} \int_{H} f^{2} d \nu,  \tag{3.27}\\
\int_{H}\|D u\|^{2} d v & \leq \frac{2}{\lambda} \int_{H} f^{2} d v, \\
\frac{1}{2} \int_{H} \operatorname{Tr}\left[\left(D^{2} u\right)^{2}\right] d v+\int_{H}\left\|Q^{-1 / 2} D u\right\|^{2} d v & \leq 4 \int_{H} f^{2} d v . \tag{3.28}
\end{align*}
$$

Proof. Let $U_{\alpha}$ be the Moreau-Yosida approximations of $U$, defined in (2.9). Since $D U_{\alpha}$ is Lipschitz continuous, we may use the results of Sections 3.2.3 and 3.2.4 for problem

$$
\begin{equation*}
\lambda u_{\alpha}-\mathcal{L} u_{\alpha}+\left\langle D U_{\alpha}, D u_{\alpha}\right\rangle=f \tag{3.29}
\end{equation*}
$$

Let $Z_{\alpha}=\int_{H} e^{-2 U_{\alpha}(x)} \mu(d x)$ and $\nu_{\alpha}:=e^{-2 U_{\alpha}} \mu / Z_{\alpha}$. Fix any $f \in \mathcal{F} \mathcal{C}_{b}^{\infty}(H), \lambda>0$, and let $u_{\alpha}$ be the strong solution to (3.29) in the space $L^{2}\left(H, v_{\alpha}\right)$. By Lemma 3.5,

$$
\begin{align*}
\int_{H} u_{\alpha}^{2} e^{-2 U_{\alpha}} d \mu & \leq \frac{1}{\lambda^{2}} \int_{H} f^{2} e^{-2 U_{\alpha}} d \mu \\
\int_{H}\left\|D u_{\alpha}\right\|^{2} e^{-2 U_{\alpha}} d \mu & \leq \frac{2}{\lambda} \int_{H} f^{2} e^{-2 U_{\alpha}} d \mu \tag{3.30}
\end{align*}
$$

and by Proposition 3.7,

$$
\begin{gather*}
\frac{1}{2} \int_{H} \operatorname{Tr}\left[\left(D^{2} u_{\alpha}\right)^{2}\right] e^{-2 U_{\alpha}} d \mu+\frac{1}{2} \int_{H}\left\|Q^{-1 / 2} D u_{\alpha}\right\|^{2} e^{-2 U_{\alpha}} d \mu \\
\quad+\int_{H}\left\langle D^{2} U_{\alpha} D u_{\alpha}, D u_{\alpha}\right\rangle e^{-2 U_{\alpha}} d \mu \leq 4 \int_{H} f^{2} e^{-2 U_{\alpha}} d \mu \tag{3.31}
\end{gather*}
$$

The right-hand sides of (3.30) and (3.31) are bounded by a constant independent of $\alpha$, since $U_{\alpha} \geq \inf U$ so that

$$
\begin{equation*}
\int_{H} f^{2} e^{-2 U_{\alpha}} d \mu \leq\|f\|_{\infty}^{2} e^{-2 \inf U} \tag{3.32}
\end{equation*}
$$

Since $U_{\alpha} \leq U$, then $e^{-2 U} \leq e^{-2 U_{\alpha}}$, and it follows that $u_{\alpha} \in W^{2,2}(H, v)$ and their $W^{2,2}(H, v)$ norms are bounded by a constant independent of $\alpha$. A sequence $\left(u_{\alpha_{n}}\right)$, with $\lim _{n \rightarrow \infty} \alpha_{n}=0$, converges weakly in $W^{2,2}(H, v)$ and in $W_{-1 / 2}^{1,2}(H, v)$ to a limit function denoted by $u$. Letting $n \rightarrow \infty$ yields that $u$ satisfies (3.27)
and (3.28). Our aim is to show that $u$ coincides with the weak solution to (1.1). For every $n$ we have

$$
\begin{aligned}
\lambda \int_{H} u_{\alpha_{n}} \varphi e^{-2 U_{\alpha_{n}}} d \mu+\frac{1}{2} \int_{H}\left\langle D u_{\alpha_{n}}, D \varphi\right\rangle e^{-2 U_{\alpha_{n}}} d \mu=\int_{H} f \varphi e^{-2 U_{\alpha_{n}}} d \mu \\
\varphi \in \mathcal{F} \mathcal{C}_{b}^{1}(H)
\end{aligned}
$$

Letting $n \rightarrow \infty$, the right-hand side converges to $\int_{H} f \varphi e^{-2 U} d \mu$. Let us split the left-hand side as

$$
\begin{aligned}
& \int_{H}\left(\lambda u_{\alpha_{n}} \varphi+\frac{1}{2}\left\langle D u_{\alpha_{n}}, D \varphi\right\rangle\right) e^{-2 U_{\alpha_{n}}} d \mu \\
& \quad=\int_{H}\left(\lambda u_{\alpha_{n}} \varphi+\frac{1}{2}\left\langle D u_{\alpha_{n}}, D \varphi\right\rangle\right) e^{-2 U} d \mu \\
& \quad+\int_{H}\left(\lambda u_{\alpha_{n}} \varphi+\frac{1}{2}\left\langle D u_{\alpha_{n}}, D \varphi\right\rangle\right)\left(1-e^{-2 U+2 U_{\alpha_{n}}}\right) e^{-2 U_{\alpha_{n}}} d \mu
\end{aligned}
$$

The first integral converges to $\int_{H}\left(\lambda u \varphi+\frac{1}{2}\langle D u, D \varphi\rangle\right) e^{-2 U} d \mu$. We claim that the second integral too vanishes as $n \rightarrow \infty$. Indeed, by the Hölder inequality with respect to the measure $e^{-2 U_{\alpha_{n}}} d \mu$, its modulus is bounded by

$$
\begin{aligned}
& \left(\int_{H}\left(\lambda u_{\alpha_{n}} \varphi+\frac{1}{2}\left\langle D u_{\alpha_{n}}, D \varphi\right\rangle\right)^{2} e^{-2 U_{\alpha_{n}}} d \mu\right)^{1 / 2} \\
& \quad \times\left(\int_{H}\left(1-e^{-2 U+2 U_{\alpha_{n}}}\right)^{2} e^{-2 U_{\alpha_{n}}} d \mu\right)^{1 / 2} \\
& \quad \leq\|\varphi\|_{C_{b}^{1}(H)}\left(\left\|\lambda u_{\alpha_{n}}\right\|_{L^{2}\left(H, e^{-2 U_{\alpha_{n}}} \mu\right)}+\frac{1}{2}\| \| D u_{\alpha_{n}}\| \|_{L^{2}\left(H, e^{-2 U_{\alpha_{n}}} \mu\right)}\right) \\
& \quad \times\left(\int_{H}\left(1-e^{-2 U+2 U_{\alpha_{n}}}\right)^{2} e^{-2 U_{\alpha_{n}}} d \mu\right)^{1 / 2} .
\end{aligned}
$$

Recalling (3.32), (3.30) implies now that

$$
\left\|\lambda u_{\alpha_{n}}\right\|_{L^{2}\left(H, e^{-2 U_{\alpha_{n}}} \mu\right)}+\frac{1}{2}\| \| D u_{\alpha_{n}}\| \|_{L^{2}\left(H, e^{-2 U_{\alpha_{n}}} \mu\right)}
$$

is bounded by a constant independent of $n$. Moreover $\int_{H}\left(1-e^{-2 U+2 U_{\alpha_{n}}}\right)^{2} \times$ $e^{-2 U_{\alpha_{n}}} d \mu$ vanishes as $n \rightarrow \infty$ by the dominated convergence theorem, and the claim is proved.

Therefore, $u$ satisfies (3.3) for every $\varphi \in \mathcal{F} \mathcal{C}_{b}^{1}(H)$, and hence it is the weak solution to (1.1).

If $f \in L^{2}(H, v)$, there is a sequence of $\mathcal{F C}_{b}^{\infty}(H)$ functions that converge to $f$ in $L^{2}(H, v)$. The sequence $\left(R(\lambda, K) f_{k}\right)$ of the weak solutions to (1.1) with $f$ replaced by $f_{k}$ converge to the weak solution $u=R(\lambda, K) f$ of (1.1), and it is a Cauchy sequence in $W^{2,2}(H, v)$ and in $W_{-1 / 2}^{1,2}(H, v)$ by estimate (3.28). Then $u \in W^{2,2}(H, v) \cap W_{-1 / 2}^{1,2}(H, v)$, and it satisfies (3.28) too.
3.4. Another maximal estimate. Under further assumptions we may recover the full estimate on $D u$ that holds in the case that $D U$ is Lipschitz continuous. In fact, we shall show below that

$$
\begin{equation*}
\int_{H}\left\langle D^{2} U D u, D u\right\rangle d v \leq 4 \int_{H} f^{2} d v \tag{3.33}
\end{equation*}
$$

in the case where $U \in C^{2}(H)$, while in Section 4.2 it will be proved in a specific example with $U \notin C^{2}(H)$. Here and in the following, we denote by $C^{2}(H)$ the space of the twice Fréchet differentiable functions from $H$ to $\mathbb{R}$, with continuous second order derivative.

We need a preliminary result.

Lemma 3.10. Under Hypothesis 2.1, for each $f \in C_{b}(H)$ there is $\alpha_{n} \rightarrow 0$ such that $u_{\alpha_{n}} \rightarrow u$ in $W^{1,2}(H, v)$ as $n \rightarrow \infty$.

Proof. We already know that there exists a sequence $\left(u_{\alpha_{n}}\right)$ weakly convergent to $u$ in $W^{1,2}(H, v)$. So, it is enough to show that

$$
\begin{equation*}
\limsup _{n \rightarrow \infty}\left|u_{\alpha_{n}}\right|_{W^{1,2}(H, v)} \leq|u|_{W^{1,2}(H, v)} \tag{3.34}
\end{equation*}
$$

for some equivalent norm $|\cdot|_{W^{1,2}(H, v)}$ in $W^{1,2}(H, v)$.
By Lemma 3.5 we have

$$
\int_{H}\left(\lambda\left|u_{\alpha_{n}}\right|^{2}+\frac{1}{2}\left\|D u_{\alpha_{n}}\right\|^{2}\right) e^{-2 U_{\alpha_{n}}} d \mu=\int_{H} f u_{\alpha_{n}} e^{-2 U_{\alpha_{n}}} d \mu .
$$

We claim that the right-hand side converges to $Z \int_{H} f u d \nu$ as $n \rightarrow \infty$. In fact we have

$$
\int_{H} f u_{\alpha_{n}} e^{-2 U_{\alpha_{n}}} d \mu=\int_{H} f u_{\alpha_{n}} e^{-2 U} d \mu+\int_{H} f u_{\alpha_{n}}\left(1-e^{2 U_{\alpha_{n}}-2 U}\right) e^{-2 U_{\alpha_{n}}} d \mu
$$

where the first addendum tends to $Z \int_{H} f u d \nu$, and the second one is estimated by

$$
\begin{aligned}
& \left|\int_{H} f u_{\alpha_{n}}\left(1-e^{2 U_{\alpha_{n}}-2 U}\right) e^{-2 U_{\alpha_{n}}} d \mu\right| \\
& \quad \leq\|f\|_{\infty}\left\|u_{\alpha_{n}}\right\|_{L^{2}\left(H, e^{-2 U_{\alpha_{n}}} \mu\right)} \int_{H}\left(1-e^{2 U_{\alpha_{n}}-2 U}\right)^{2} e^{-2 U_{\alpha_{n}}} d \mu
\end{aligned}
$$

which vanishes as $n \rightarrow \infty$ because $\left\|u_{\alpha_{n}}\right\|_{L^{2}\left(H, e^{-2 U_{\alpha_{n}}} \mu\right)}$ is bounded and

$$
\lim _{n \rightarrow \infty} \int_{H}\left(1-e^{2 U_{\alpha_{n}}-2 U}\right)^{2} e^{-2 U_{\alpha_{n}}} d \mu=0
$$

by the dominated convergence theorem.

Therefore we have

$$
\begin{aligned}
& \limsup _{n \rightarrow \infty} \int_{H}\left(\lambda u_{\alpha_{n}}^{2}+\frac{1}{2}\left\|D u_{\alpha_{n}}\right\|^{2}\right) e^{-2 U} d \mu \\
& \quad \leq \limsup _{n \rightarrow \infty} \int_{H}\left(\lambda\left|u_{\alpha_{n}}\right|^{2}+\frac{1}{2}\left\|D u_{\alpha_{n}}\right\|^{2}\right) e^{-2 U_{\alpha_{n}}} d \mu=Z \int_{H} f u d \nu
\end{aligned}
$$

Moreover

$$
\int_{H} f u d v=\int_{H}\left(\lambda u^{2}+\frac{1}{2}\|D u\|^{2}\right) d v
$$

so that

$$
\limsup _{n \rightarrow \infty} \int_{H}\left(\lambda\left|u_{\alpha_{n}}\right|^{2}+\frac{1}{2}\left\|D u_{\alpha_{n}}\right\|^{2}\right) d v \leq \int_{H}\left(\lambda u^{2}+\frac{1}{2}\|D u\|^{2}\right) d \nu
$$

and (3.34) follows.
Now we can prove estimate (3.33).
THEOREM 3.11. Let $U$ be a $C^{2}$ function satisfying Hypothesis 2.1. Then (3.33) is fulfilled for all $f \in L^{2}(H, v)$.

Proof. Since $C_{b}(H)$ is dense in $L^{2}(H, v)$ it is sufficient to prove (3.33) when $f \in C_{b}(H)$. In this case, let $\alpha_{n} \rightarrow 0$ be such that $u_{\alpha_{n}} \rightarrow u$ in $W^{1,2}(H, v)$ (Lemma 3.10). Then $D u_{\alpha_{n}} \rightarrow D u$ in $L^{2}(H, v ; H)$ and so [possibly replacing $\left(\alpha_{n}\right)$ by a subsequence] $D u_{\alpha_{n}}(x) \rightarrow D u(x)$ for almost all $x$. Using Lemma 2.11, for these $x$ we have
$\lim _{n \rightarrow \infty}\left\langle D^{2} U_{\alpha_{n}}(x) D u_{\alpha_{n}}(x), D u_{\alpha_{n}}(x)\right\rangle e^{-2 U_{\alpha_{n}}(x)}=\left\langle D^{2} U(x) D u(x), D u(x)\right\rangle e^{-2 U(x)}$, and by Fatou's lemma,

$$
\begin{aligned}
\int_{H}\langle & \left.D^{2} U(x) D u(x), D(x)\right\rangle d v \\
& =\int_{H}\left\langle D^{2} U(x) D u(x), D(x)\right) e^{-2 U(x)} d \mu \\
& \leq \liminf _{n \rightarrow \infty} \int_{H}\left\langle D^{2} U_{\alpha_{n}}(x) D u_{\alpha_{n}}(x), D u_{\alpha_{n}}(x)\right) e^{-2 U_{\alpha_{n}}(x)} d \mu \\
& \leq 4 \liminf _{n \rightarrow \infty} \int_{H} f^{2} e^{-2 U_{\alpha_{n}}} d \mu=4 \int_{H} f^{2} d \nu
\end{aligned}
$$

4. Perturbations. The regularity results and estimates of Section 3 open the way to new results for nonsymmetric Kolmogorov operators, by perturbation. Here we consider the operator $K_{1}$ in the space $L^{2}(H, v)$ defined by

$$
\begin{equation*}
D\left(K_{1}\right)=D(K), \quad K_{1} v:=K v+\langle B(x), D v(x)\rangle \tag{4.1}
\end{equation*}
$$

with a (possibly) nongradient field $B: H \mapsto H$.
We shall give two perturbation results, the first one in the general case (Section 4.1) and the second one in the case where the weak solution to (1.1) satisfies (3.33) (Section 4.2). In both cases we shall use the next proposition and a part of its proof.

Proposition 4.1. Let A be a self-adjoint dissipative operator in $L^{2}(H, v)$, and let $\mathcal{B}: D(A) \mapsto L^{2}(H, v)$ be a linear operator such that

$$
\begin{equation*}
\|\mathcal{B} v\|_{L^{2}(H, v)}^{2} \leq a\|A v\|_{L^{2}(H, v)}^{2}+b\|v\|_{L^{2}(H, v)}^{2}, \quad v \in D(A), \tag{4.2}
\end{equation*}
$$

for some $a<1 /(\sqrt{2}+1)^{2}$ and $b>0$. Then the operator

$$
A_{1}: D(A) \mapsto L^{2}(H, v), \quad A_{1} v=A v+\mathcal{B} v
$$

generates an analytic semigroup in $L^{2}(H, v)$.
Proof. Let us denote by $\mathcal{X}=L^{2}(H, v ; \mathbb{C})$ the complexification of $L^{2}(H, v)$ and by $\mathcal{A}$ the complexification of $A, \mathcal{A}(u+i v)=A u+i A v$. Then the spectrum of $\mathcal{A}$ is contained in $(-\infty, 0]$, and we have $\|\lambda R(\lambda, \mathcal{A})\|_{\mathcal{L}(\mathcal{X})} \leq 1 / \cos (\theta / 2)$ for $\lambda \in$ $\mathbb{C} \backslash(-\infty, 0]$, with $\theta=\arg \lambda$. Hence, for $\operatorname{Re} \lambda>0$ we have $\|\lambda R(\lambda, \mathcal{A})\|_{\mathcal{L}(\mathcal{X})} \leq \sqrt{2}$.

A standard general perturbation result for analytic semigroups in Banach spaces states that if the generator $\mathcal{A}$ of an analytic semigroup in a complex Banach space $\mathcal{X}$ satisfies $\|\lambda R(\lambda, \mathcal{A})\|_{\mathcal{L}(\mathcal{X})} \leq M$ for $\operatorname{Re} \lambda>\omega$, then for any linear perturbation $\mathcal{B}: D(\mathcal{A}) \mapsto \mathcal{X}$ that satisfies

$$
\|\mathcal{B} v\|_{\mathcal{X}} \leq c_{1}\|A v\|_{\mathcal{X}}+c_{2}\|v\|_{\mathcal{X}}, \quad v \in D(\mathcal{A})
$$

with $c_{1}<1 /(M+1)$ and $c_{2} \in \mathbb{R}$, the sum $\mathcal{A}+\mathcal{B}: D(\mathcal{A}) \mapsto \mathcal{X}$ generates an analytic semigroup in $\mathcal{X}$. We write down a proof, which will be used later.

For $\operatorname{Re} \lambda>\omega$ the resolvent equation $\lambda u-(\mathcal{A}+\mathcal{B}) u=f$ is equivalent (setting $\lambda u-\mathcal{A} u=v$ ) to the fixed point problem $v=T v$, with $T: \mathcal{X} \mapsto \mathcal{X}, T v=$ $\mathcal{B} R(\lambda, \mathcal{A}) v+f$. We have

$$
\begin{aligned}
\|T v\| & \leq c_{1}\|\mathcal{A} R(\lambda, A) v\|+c_{2}\|R(\lambda, \mathcal{A}) v\| \\
& \leq c_{1}(M+1)\|v\|+\frac{c_{2} M}{|\lambda|}\|v\|, \quad v \in \mathcal{X} .
\end{aligned}
$$

Fix $\omega_{0}>\omega$ such that $C:=c_{1}(M+1)+c_{2} M / \omega_{0}<1$. Then for every $\lambda$ in the halfplane $\operatorname{Re} \lambda \geq \omega_{0} T$ is a contraction with constant $C$, the equation $v=T v$ has a unique solution $v \in \mathcal{X}$ and $\|v\| \leq\|f\| /(1-C)$, and the resolvent equation $\lambda u-$ $A_{1} u=f$ has a unique solution $u=R(\lambda, A) v$ with $\|u\| \leq M\|f\| /|\lambda|(1-C)$, and the statement follows.

In our case we can take $\omega=0$ and $M=\sqrt{2}$. Assumption (4.2) implies that $\|\mathcal{B} v\|_{\mathcal{X}} \leq \sqrt{a}\|A v\|_{\mathcal{X}}+\sqrt{b}\|v\|_{\mathcal{X}}$, for every $v \in D(\mathcal{A})$, so we require $a<$ $1 /(\sqrt{2}+1)^{2}$. Once we know that $\mathcal{A}+\mathcal{B}$ generates an analytic semigroup $T(t)$ in $L^{2}(H, v ; \mathbb{C})$, it is sufficient to remark that the restriction of $T(t)$ to $L^{2}(H, v)$ preserves $L^{2}(H, v)$, and it is an analytic semigroup in $L^{2}(H, v)$.

### 4.1. First perturbation.

Proposition 4.2. Let $U$ satisfy Hypothesis 2.1. Let $B: H \mapsto H$ be $\mu$-measurable (hence, $v$-measurable) and such that there exist $c_{1} \in(0,1 / 2(\sqrt{2}+1))$, $c_{2}>0$ such that for a.e. $x \in H$ we have

$$
\begin{equation*}
|\langle B(x), y\rangle| \leq c_{1}\left\|Q^{-1 / 2} y\right\|+c_{2}\|y\|, \quad y \in Q^{1 / 2}(H) \tag{4.3}
\end{equation*}
$$

Then the operator $K_{1}$ defined in (4.1) generates an analytic semigroup in $L^{2}(H, v)$. In particular, there exist $\lambda_{0} \geq 0, C>0$ such that for every $\lambda>\lambda_{0}$ and for every $f \in L^{2}(H, v)$ the equation $\lambda v-K_{1} v=f$ has a unique solution $v \in D(K)$, and

$$
\|v\|_{D(K)} \leq C\|f\|_{L^{2}(H, v)} .
$$

Proof. In view of Proposition 4.1, it is sufficient to show that the operator $\mathcal{B}$ defined in $D(K)$ by

$$
\mathcal{B} u(x)=\langle B(x), D u(x)\rangle, \quad x \in H,
$$

satisfies estimate

$$
\begin{equation*}
\|\mathcal{B} v\|_{L^{2}(H, v)}^{2} \leq a\|K v\|_{L^{2}(H, v)}^{2}+b\|v\|_{L^{2}(H, v)}^{2}, \quad v \in D(K) \tag{4.4}
\end{equation*}
$$

for some $a<(\sqrt{2}+1)^{-2}$. We note that for every $u \in D(K)$ we have

$$
\begin{align*}
\int_{H}\|D u\|^{2} d v & \leq 4 \lambda \int_{H} u^{2} d v+\frac{4}{\lambda} \int_{H}(K u)^{2} d v \quad \forall \lambda>0  \tag{4.5}\\
\int_{H}\left\|Q^{-1 / 2} D u\right\|^{2} d v & \leq 4 \int_{H}(K u)^{2} d v \tag{4.6}
\end{align*}
$$

Estimate (4.5) follows from (3.27), taking $f=\lambda u-K u$. Estimate (4.6) follows from (3.28) taking again $f=\lambda u-K u$, and letting $\lambda \rightarrow 0$. Using (4.5) and (4.6), for each $\varepsilon \in(0,1)$ and $\lambda>0$ we get

$$
\begin{aligned}
\int_{H}\langle B, D u\rangle^{2} d v \leq & \int_{H}\left(c_{1}\left\|Q^{-1 / 2} D u\right\|+c_{2}\|D u\|\right)^{2} d v \\
\leq & c_{1}^{2}(1+\varepsilon) \int_{H}\left\|Q^{-1 / 2} D u\right\|^{2} d v+c_{2}^{2}\left(1+\frac{1}{\varepsilon}\right) \int_{H}\|D u\|^{2} d v \\
\leq & 4 c_{1}^{2}(1+\varepsilon) \int_{H}(K u)^{2} d v \\
& +c_{2}^{2}\left(1+\frac{1}{\varepsilon}\right)\left(4 \lambda \int_{H} u^{2} d v+\frac{4}{\lambda} \int_{H}(K u)^{2} d v\right)
\end{aligned}
$$

Since $4 c_{1}^{2}<1 /(\sqrt{2}+1)^{2}$, there is $\varepsilon>0$ such that $4 c_{1}^{2}(1+\varepsilon)<1 /(\sqrt{2}+1)^{2}$. Fixed such $\varepsilon$, choose $\lambda$ big enough, such that $a:=4 c_{1}^{2}(1+\varepsilon)+4 c_{2}^{2}(1+1 / \varepsilon) / \lambda<$ $1 /(\sqrt{2}+1)^{2}$. With these choices estimate (4.4) is satisfied with $a<1 /(\sqrt{2}+1)^{2}$, and the statement follows from Proposition 4.1.

REMARK 4.3. The assumptions of Proposition 4.2 are satisfied if $x \mapsto$ $Q^{\alpha} B(x) \in L^{\infty}(H, v ; H)$ for some $\alpha<1 / 2$. Indeed, in this case for $y \in Q^{1 / 2}(H)$ and a.e. $x \in H$, we have

$$
\begin{array}{r}
|\langle B(x), y\rangle|=\left|\left\langle Q^{\alpha} B(x), Q^{-\alpha} y\right\rangle\right| \leq\left\|Q^{\alpha} B(\cdot)\right\|_{\infty}\left(\varepsilon\left\|Q^{-1 / 2} y\right\|+c(\varepsilon)\|y\|\right) \\
x \in H, \varepsilon>0
\end{array}
$$

and choosing $\varepsilon$ small enough, (4.3) is satisfied with $c_{1}<1 / 2(\sqrt{2}-1)$.
In the case that $x \mapsto Q^{1 / 2} B(x) \in L^{\infty}(H, v ; H)$ we need some restriction in order that the assumptions of Proposition 4.2 be satisfied. For instance, they are satisfied if $B=B_{1}+B_{2}$, with $B_{1} \in L^{\infty}(H, v ; H)$ and $Q^{1 / 2} B_{2} \in L^{\infty}(H, v ; H)$, $\left\|Q^{1 / 2} B_{2}\right\|_{\infty} \leq c_{1}<1 / 2(\sqrt{2}+1)$.
4.2. Second perturbation. In the case that $U \in C^{2}(H)$ we have also estimate (3.33), which is useful when

$$
\begin{equation*}
\left\langle D^{2} U(x) y, y\right\rangle \geq C(x)\|y\|^{2}, \quad x, y \in H \tag{4.7}
\end{equation*}
$$

and the function $C(x)$ is unbounded from above [if $C$ is bounded from above, (3.33) does not add much information to (3.27)].

Proposition 4.4. Let $U \in C^{2}(H)$ satisfy Hypothesis 2.1. Assume moreover that (4.7) holds for some unbounded $C(x)$ and that for every $\lambda>0$ and $f \in$ $L^{2}(H, v)$ the weak solution $u$ to (1.1) satisfies (3.33). Moreover, let $B: H \mapsto H$ be $\mu$-measurable and such that there exist $c_{1}, c_{2}, c_{3}>0$ with $c_{1}^{2}+c_{2}^{2}<1 / 8(\sqrt{2}+1)^{2}$, and for a.e. $x \in H$, we have

$$
\begin{equation*}
|\langle B(x), y\rangle| \leq c_{1}\left\|Q^{-1 / 2} y\right\|+c_{2} \sqrt{C(x)}\|y\|+c_{3}\|y\|, \quad y \in Q^{1 / 2}(H) \tag{4.8}
\end{equation*}
$$

Then the operator $K_{1}$ defined in (4.1) generates an analytic semigroup in $L^{2}(H, v)$. In particular, there exist $\lambda_{0} \geq 0, C>0$ such that for every $\lambda>\lambda_{0}$ and for every $f \in L^{2}(H, v)$ the equation $\lambda v-K_{1} v=f$ has a unique solution $v \in D(K)$, and

$$
\|v\|_{D(K)} \leq C\|f\|_{L^{2}(H, v)} .
$$

Proof. We argue as in the proof of Proposition 4.2. Here, besides estimates (4.5) and (4.6), we also use

$$
\begin{equation*}
\int_{H}\left\langle D^{2} U D u, D u\right\rangle d v \leq 4 \int_{H}(K u)^{2} d v, \quad u \in D(K) \tag{4.9}
\end{equation*}
$$

which follows from (3.33) taking $f=\lambda u-K u$ and letting $\lambda \rightarrow 0$. By (4.8) for each $u \in D(K)$ we have

$$
\int_{H}\langle B, D u\rangle^{2} d v \leq \int_{H}\left(c_{1}\left\|Q^{-1 / 2} D u\right\|+c_{2} \sqrt{C(x)}\|D u\|+c_{3}\|D u\|\right)^{2} d \nu
$$

Using the inequalities $(a+b+c)^{2} \leq a^{2}(2+\varepsilon)+b^{2}(2+\varepsilon)+c^{2}(1+2 / \varepsilon)$ for each $\varepsilon \in(0,1)$, and

$$
\int_{H} C(x)\|D u\|^{2} d v \leq \int_{H}\left\langle D^{2} U D u, D u\right\rangle d v \leq 4 \int_{H}(K u)^{2} d v
$$

that follows from (4.7) and (4.9), we obtain, recalling (4.5) and (4.6),

$$
\begin{aligned}
& \int_{H}\langle B, D u\rangle^{2} d v \\
& \leq c_{1}^{2}(2+\varepsilon) \int_{H}\left\|Q^{-1 / 2} D u\right\|^{2} d v+c_{2}^{2}(2+\varepsilon) \int_{H} C(x)\|D u\|^{2} d v \\
&+c_{3}^{2}\left(1+\frac{2}{\varepsilon}\right) \int_{H}\|D u\|^{2} d v \\
& \leq 4\left(c_{1}^{2}+c_{2}^{2}\right)(2+\varepsilon) \int_{H}(K u)^{2} d v \\
&+c_{3}^{2}\left(1+\frac{2}{\varepsilon}\right)\left(4 \lambda \int_{H} u^{2} d v+\frac{4}{\lambda} \int_{H}(K u)^{2} d v\right)
\end{aligned}
$$

As in the proof of Proposition 4.2, we may choose $\varepsilon$ small and then $\lambda$ large, in such a way that for every $u \in D(K)$, we have $\int_{H}\langle B, D u\rangle^{2} d v \leq a \int_{H}(K u)^{2} d v+$ $b \int_{H} u^{2} d v$ with $a<1 /(\sqrt{2}+1)^{2}$, and the statement follows from Proposition 4.1.

REMARK 4.5. Assumption (4.8) is satisfied if $B=B_{1}+B_{2}$, where $x \mapsto$ $Q^{\alpha} B_{1}(x) \in L^{\infty}(H, v ; H)$ for some $\alpha \in[1 / 2)$ and there are $b<1 / 2(2+\sqrt{2})$, $c>0$ such that $\left\|B_{2}(x)\right\| \leq b C(x)+c$ for almost every $x \in H$.

Theorem 3.11 allows to use Proposition 4.4 when $U \in C^{2}(H)$. In some specific examples the result of Proposition 4.4 holds when $U$ is not $C^{2}$, but belongs to a suitable Sobolev space. See Section 5.2.

We emphasize that the domain of the perturbed operator $K_{1}$ coincides with $D(K)$. Therefore, under the assumptions of Proposition 4.2 for every $u \in D\left(K_{1}\right)$ we have

$$
u \in W^{2,2}(H, v), \quad \int_{H}\left\|A^{-1 / 2} D u\right\|^{2} d v<\infty
$$

and if the assumptions of Proposition 4.4 hold, then for every $u \in D\left(K_{1}\right)$ we have also

$$
\int_{H}\left\langle D^{2} U D u, D u\right\rangle d v<\infty
$$

An important feature of the semigroup generated by $K_{1}$ is positivity preserving. If $B \equiv 0$, that is $K_{1}=K$, Lemma 2.7 implies that $K$ satisfies the Beurling-Deny conditions that yield positivity preserving (e.g., [6], Sections 1.3, 1.4).

Proposition 4.6. Let the assumptions of Proposition 4.2 or of Proposition 4.4 hold, and let $\lambda_{0}$ be given by Proposition 4.2 or 4.4. Then for every $\lambda>\lambda_{0}$ and $f \in L^{2}(H, v)$ such that $f(x) \geq 0$ a.e., $R\left(\lambda, K_{1}\right) f(x) \geq 0$ a.e.

Proof. Let us introduce the approximations

$$
B_{n}(x):=n R(n, A) B(x) \mathbb{1}_{\{x \in H:\|B(x)\| \leq n\}}, \quad n \in \mathbb{N}, x \in H
$$

that are $\mu$-measurable and bounded in $H$.
If the assumptions of Proposition 4.2 hold, then each $B_{n}$ satisfies (4.2) with the same constants $a, b$ of $B$. Indeed, since $\|n R(n, A)\|_{\mathcal{L}(H)} \leq 1$, then for every $x \in H$ and $y \in Q^{1 / 2}(H)$ we have

$$
\begin{aligned}
\left|\left\langle B_{n}(x), y\right\rangle\right| & =|\langle B(x), n R(n, A) y\rangle| \mathbb{1}_{\{x \in H:\|B(x)\| \leq n\}} \\
& \leq a\left\|Q^{-1 / 2} n R(n, A) y\right\|+b\|n R(n, A) y\| \\
& =a\left\|n R(n, A) Q^{-1 / 2} y\right\|+b\|n R(n, A) y\| \leq a\left\|Q^{-1 / 2} y\right\|+b\|y\| .
\end{aligned}
$$

Similarly, if the assumptions of Proposition 4.4 hold, then $B_{n}$ satisfies (4.8) with the same constants $c_{1}, c_{2}, c_{3}$ as $B$. Moreover $B_{n}$ converges to $B v$-a.e., since

$$
B_{n}(x)-B(x)=n R(n, A) B(x)-B(x) \quad \text { if }\|B(x)\| \leq n .
$$

For each $f \in L^{2}(H, v)$ we may approach $R\left(\lambda, K_{1}\right) f$ by the solutions $u_{n} \in$ $D(K)$ of problems

$$
\begin{equation*}
\lambda u_{n}-K u_{n}-\left\langle B_{n}(x), D u_{n}\right\rangle=f \tag{4.10}
\end{equation*}
$$

that still exist for $\lambda>\lambda_{0}$ since the functions $B_{n}$ satisfy the assumptions of Proposition 4.1 (or, of Proposition 4.4) with the same constants as $B$. By the proof of Propositions 4.2 and 4.4, $u_{n}$ is obtained as $R(\lambda, K)\left(I-T_{n}\right)^{-1}$ where

$$
T_{n} v=\left\langle B_{n}(\cdot), D R(\lambda, K) v\right\rangle, \quad v \in L^{2}(H, v)
$$

and $\left(I-T_{n}\right)^{-1}$ exists because $T$ is a contraction. We may use the principle of contractions depending on a parameter since

$$
\left\|T_{n} v-T v\right\|_{L^{2}(H, v)}^{2} \leq \int_{H}\left|\left\langle B-B_{n}, D R(\lambda, K) v\right\rangle\right|^{2} d v
$$

that vanishes as $n \rightarrow \infty$ by the dominated convergence theorem. Indeed, for $v$ almost every $x$ we have $\lim _{n \rightarrow \infty} B_{n}(x)=B(x)$ and

$$
\left|\left\langle B_{n}(x), D R(\lambda, K) v(x)\right\rangle\right| \leq a\left\|Q^{-1 / 2} D R(\lambda, K) v(x)\right\|+b\|D R(\lambda, K) v(x)\|
$$

if the assumptions of Proposition 4.2 hold, and

$$
\begin{aligned}
\left|\left\langle B_{n}(x), D R(\lambda, K) v(x)\right\rangle\right| \leq & c_{1}\left\|Q^{-1 / 2} D R(\lambda, K) v(x)\right\| \\
& +c_{2} \sqrt{C(x)}\|D R(\lambda, K) v(x)\| \\
& +c_{3}\|D R(\lambda, K) v(x)\|
\end{aligned}
$$

if the assumptions of Proposition 4.4 hold. In both cases, the right-hand sides belong to $L^{2}(H, v)$.

It follows that for $\lambda>\lambda_{0}$ we have $\lim _{n \rightarrow \infty} u_{n}=R\left(\lambda, K_{1}\right) f$, in $L^{2}(H, v)$. To finish the proof we show that if $f \geq 0 v$-a.e., then $u_{n} \geq 0 v$-a.e. This will yield the statement.

Let us multiply both sides of (4.10) by $u_{n}^{-}$, that belongs to $W^{1,2}(H, v)$ by Lemma 2.7, and integrate over $H$. We get

$$
\lambda \int_{H} u_{n} u_{n}^{-} d v+\frac{1}{2} \int_{H}\left\langle D u_{n}, D u_{n}^{-}\right\rangle d v-\int_{H}\left\langle B_{n}, D u_{n}\right\rangle u_{n}^{-} d v=\int_{H} f u_{n}^{-} d v,
$$

and recalling that $u_{n} u_{n}^{-}=-\left(u_{n}^{-}\right)^{2},\left\langle D u_{n}, D u_{n}^{-}\right\rangle=-\left\|D u_{n}^{-}\right\|^{2}$ by Lemma 2.7, we obtain

$$
-\lambda \int_{H}\left(u_{n}^{-}\right)^{2} d v-\frac{1}{2} \int_{H}\left\|D u_{n}^{-}\right\|^{2} d v-\int_{H}\left\langle B_{n}, D u_{n}\right\rangle u_{n}^{-} d v \geq 0 .
$$

Now we estimate

$$
\begin{aligned}
\left|\int_{H}\left\langle B_{n}, D u_{n}\right\rangle u_{n}^{-} d \nu\right| & =\left|\int_{\left\{u_{n} \leq 0\right\}}\left\langle B_{n}, D u_{n}\right\rangle u_{n}^{-} d v\right| \\
& =\mid \int_{H}\left\langle B_{n}, D u_{n}^{-}\right| u_{n}^{-} d \nu \mid \\
& \leq\left\|B_{n}\right\|_{\infty}\left(\int_{H}\left\|D u_{n}^{-}\right\|^{2} d v\right)^{1 / 2}\left(\int_{H}\left(u_{n}^{-}\right)^{2} d v\right)^{1 / 2} \\
& \leq \frac{1}{2} \int_{H}\left\|D u_{n}^{-}\right\|^{2} d v+2\left\|B_{n}\right\|_{\infty} \int_{H}\left(u_{n}^{-}\right)^{2} d v
\end{aligned}
$$

If $\lambda>C_{n}:=2\left\|B_{n}\right\|_{\infty}$, we get

$$
-\left(\lambda-C_{n}\right)\left\|u_{n}^{-}\right\|_{L^{2}(H, v)}^{2} \geq 0
$$

which implies $u_{n}^{-} \equiv 0$, namely $u_{n} \geq 0$ a.e. So, the resolvent of $K_{n}:=K+\left\langle B_{n}, D \cdot\right\rangle$ preserves positivity for $\lambda$ large, possibly depending on $n$. Since $K_{n}$ generates a $C_{0}$ semigroup, its resolvent preserves positivity for every $\lambda$ bigger than the type of the semigroup, in particular for every $\lambda>\lambda_{0}$. Then, $R\left(\lambda, K_{1}\right)$ preserves positivity for $\lambda>\lambda_{0}$.

Now we discuss the existence of an invariant measure $\zeta(d x)=\rho(x) \nu(d x)$ for the semigroup generated by $K_{1}$ in $L^{2}(H, v)$. An important step is the following proposition.

Proposition 4.7. Let the assumptions of Proposition 4.2 or of Proposition 4.4 hold. Let in addition Hypothesis 2.8 hold. Then the kernel of $K_{1}^{*}$ [the adjoint of $K_{1}$ in $\left.L^{2}(H, v)\right]$ contains a nonnegative function $\rho \not \equiv 0$.

Proof. The function $\mathbb{1}$ identically equal to 1 belongs to the domain of $K_{1}$, and $K_{1} \mathbb{1}=0$. Then for any $\lambda>\lambda_{0}$, $\mathbb{1}$ is an eigenvector of $R\left(\lambda, K_{1}\right)$ with eigenvalue $1 / \lambda$. Since $D\left(K_{1}\right)=D(K)$ is compactly embedded in $L^{2}(H, v)$ by Proposition 2.10, then $R\left(\lambda, K_{1}\right)$ is a compact operator, and $1 / \lambda$ is an eigenvalue of $R\left(\lambda, K_{1}\right)^{*}=R\left(\lambda, K_{1}^{*}\right)$ too. Hence, 0 is an eigenvalue of $K_{1}^{*}$, so that the kernel of $K_{1}^{*}$ contains nonzero elements. Note that since $R\left(\lambda, K_{1}\right)$ preserves positivity for large $\lambda$, then $R\left(\lambda, K_{1}^{*}\right)$ too preserves positivity for large $\lambda$, hence the semigroup $e^{t K_{1}^{*}}$ generated by $K_{1}^{*}$ preserves positivity for every $t>0$.

Let us check that the kernel of $K_{1}^{*}$ is a lattice, that is, if $\varphi \in \operatorname{Ker} K_{1}^{*}$, then $|\varphi| \in$ $\operatorname{Ker} K_{1}^{*}$. Assume that $\varphi \in \operatorname{Ker} K_{1}^{*}$. Then $\varphi=e^{t K_{1}^{*}} \varphi$ for every $t>0$, and since $e^{t K_{1}^{*}}$ preserves positivity, then

$$
|\varphi(x)|=\left|e^{t K_{1}^{*}} \varphi(x)\right| \leq\left(e^{t K_{1}^{*}}|\varphi|\right)(x), \quad \text { v-a.e. } x \in H .
$$

We claim that for every $t>0$,

$$
\begin{equation*}
|\varphi(x)|=e^{t K_{1}^{*}}(|\varphi|)(x), \quad \nu \text {-a.e. } x \in H . \tag{4.11}
\end{equation*}
$$

Assume by contradiction that there are $t>0$ and a Borel subset $I \subset H$ such that $\nu(I)>0$ and $|\varphi(x)|<e^{t K_{1}^{*}}(|\varphi|)(x)$ for $x \in I$. Then we have

$$
\int_{H}|\varphi(x)| v(d x)<\int_{H}\left(e^{t K_{1}^{*}}|\varphi|\right)(x) v(d x) .
$$

On the other hand, since $\mathbb{1} \in \operatorname{Ker} K_{1}$, then $e^{t K_{1}^{*}} \mathbb{1}=\mathbb{1}$. Hence

$$
\int_{H} e^{t K_{1}^{*}}|\varphi| d \nu=\left\langle e^{t K_{1}^{*}}\right| \varphi|, \mathbb{1}\rangle_{L^{2}(H, \nu)}=\langle | \varphi\left|, e^{t K_{1}} \mathbb{1}\right\rangle_{L^{2}(H, \nu)}=\int_{H}|\varphi| d \nu,
$$

which is a contradiction. Then (4.11) holds and it yields $|\varphi| \in \operatorname{Ker} K_{1}^{*}$.
A realization of $\mathcal{K}_{1}$ in $L^{2}(H, \rho v)$ is $m$-dissipative, as the next proposition shows.

Proposition 4.8. Under the assumptions of Proposition 4.7, let $\rho$ be a nonnegative function belonging to $\operatorname{Ker} K_{1}^{*} \backslash\{0\}$. Then the operator
$\mathcal{D}:=\left\{u \in D\left(K_{1}\right) \cap L^{2}(H, \rho \nu): K_{1} u \in L^{2}(H, \rho v)\right\} \mapsto L^{2}(H, \rho \nu), \quad u \mapsto K_{1} u$ is dissipative in $L^{2}(H, \rho \nu)$ and the range of $\lambda I-K_{1}: \mathcal{D} \mapsto L^{2}(H, \rho \nu)$ is dense in $L^{2}(H, \rho \nu)$ for $\lambda>0$. Then its closure $\widetilde{K}_{1}$ generates a contraction semigroup $\widetilde{T}_{1}(t)$ in $L^{2}(H, \rho \nu)$, and the measure $\rho v$ is invariant for $\widetilde{T}_{1}(t)$.

Proof. As a first step we prove dissipativity, through estimates on $R\left(\lambda, K_{1}\right)$.
We remark that Lemma 2.2 holds for the measure $\rho v$ as well, with the same proof. In particular, $C_{b}(H)$ is dense in $L^{1}(H, \rho \nu)$.

Let $\lambda>\lambda_{0}$ and let $f \in C_{b}(H)$. Set $u=R\left(\lambda, K_{1}\right) f$. We recall that, since $\rho \in D\left(K_{1}^{*}\right)$ and $K_{1}^{*} \rho=0$, then for every $u \in D\left(K_{1}\right)$ we have $\int_{H} K_{1} u \rho d v=$
$\int_{H} u K_{1}^{*} \rho d \nu=0$. So, multiplying both sides of $\lambda u-K_{1} u=f$ by $\rho$ and integrating we obtain

$$
\int_{H} \lambda u \rho d v=\int_{H} f \rho d v
$$

If $f$ has nonnegative values $v$-a.e., by Proposition $4.6 u$ has nonnegative values $v$-a.e., and the above equality implies

$$
\begin{equation*}
\|u\|_{L^{1}(H, \rho \nu)} \leq \frac{1}{\lambda}\|f\|_{L^{1}(H, \rho \nu)} . \tag{4.12}
\end{equation*}
$$

In general, we split $f$ as $f=f^{+}-f^{-}$. Since $u=R\left(\lambda, K_{1}\right) f^{+}-R\left(\lambda, K_{1}\right) f^{-}=$ $u^{+}-u^{-}$, (4.12) follows for every $f \in C_{b}(H)$. Since $C_{b}(H)$ is dense in $L^{1}(H, \rho v)$, the resolvent $R\left(\lambda, K_{1}\right)$ may be extended to a bounded operator [still denoted by $\left.R\left(\lambda, K_{1}\right)\right]$ to $L^{1}(H, \rho \nu)$, and

$$
\begin{equation*}
\left\|R\left(\lambda, K_{1}\right) f\right\|_{L^{1}(H, \rho \nu)} \leq \frac{1}{\lambda}\|f\|_{L^{1}(H, \rho \nu)}, \quad f \in L^{1}(H, \rho \nu) . \tag{4.13}
\end{equation*}
$$

Let now $f \in L^{\infty}(H, \rho \nu) . f$ is in fact an equivalence class of functions, that contains a Borel bounded element. Indeed, for each element $\varphi \in f$, setting $\tilde{f}(x)=$ $\varphi(x)$ if $|\varphi(x)| \leq\|f\|_{L^{\infty}(H, \rho \nu)}, \tilde{f}(x)=0$ if $|\varphi(x)|>\|f\|_{L^{\infty}(H, \rho \nu)}$, the function $\tilde{f}$ is Borel and bounded, and $\|f\|_{L^{\infty}(H, \rho \nu)}=\sup _{x \in H}|\tilde{f}(x)|$.

Let us go back to the resolvent equation, $\lambda u-K_{1} u=\tilde{f}$. Since $\tilde{f}$ is Borel and bounded, it can be seen as an element of $L^{\infty}(H, v)$, identifying it with its equivalence class. ${ }^{1}$ Moreover, $\|\tilde{f}\|_{L^{\infty}(H, v)}=\sup _{x \in H}|\tilde{f}(x)|=\|\tilde{f}\|_{L^{\infty}(H, \rho \nu)}$.

Since $\sup |\tilde{f}|-\tilde{f}(x) \geq 0$ for every $x$, still by Proposition 4.6 we have $R\left(\lambda, K_{1}\right)(\sup |\tilde{f}|-\tilde{f})=\sup |\tilde{f}| / \lambda-\underset{\tilde{f}}{u} \geq 0, v$-a.e. Similarly, since $\tilde{f}(x)+$ $\sup |\tilde{f}| \geq 0$ for every $x$, then $u+\sup |\tilde{f}| / \lambda \geq 0$, $v$-a.e. So, we get an $L^{\infty}$ estimate, $\|u\|_{L^{\infty}(H, v)} \leq \sup |\tilde{f}| / \lambda$. Hence

$$
\begin{align*}
&\left\|R\left(\lambda, K_{1}\right) f\right\|_{L^{\infty}(H, \rho v)} \leq\left\|R\left(\lambda, K_{1}\right) \tilde{f}\right\|_{L^{\infty}(H, v)} \leq \frac{1}{\lambda}\|f\|_{L^{\infty}(H, \rho \nu)}  \tag{4.14}\\
& f \in L^{\infty}(H, \rho v)
\end{align*}
$$

By interpolation, $R\left(\lambda, K_{1}\right)$ may be extended to $L^{2}(H, \rho \nu)$ [and, in fact, to all spaces $L^{p}(H, \rho \nu)$ ], in such a way that the norm of the extension does not exceed $1 / \lambda$. In particular,

$$
\begin{align*}
\left\|R\left(\lambda, K_{1}\right) f\right\|_{L^{2}(H, \rho v)} \leq \frac{1}{\lambda}\|f\|_{L^{2}(H, \rho v)} &  \tag{4.15}\\
& f \in L^{2}(H, \rho v) \cap L^{2}(H, v) .
\end{align*}
$$

[^0]Let now $u \in \mathcal{D}$. For $\lambda>\lambda_{0}$ estimate (4.15) gives

$$
\lambda\|u\|_{L^{2}(H, \rho v)} \leq\left\|\lambda u-K_{1} u\right\|_{L^{2}(H, \rho v)}
$$

and squaring the norms of both sides, we obtain

$$
\left\langle u, K_{1} u\right\rangle_{L^{2}(H, \rho v)} \leq \frac{1}{2 \lambda}\left\|K_{1} u\right\|_{L^{2}(H, \rho v)}^{2} .
$$

Letting $\lambda \rightarrow \infty$ yields $\left\langle u, K_{1} u\right\rangle_{L^{2}(H, \rho v)} \leq 0$, namely the restriction of $K_{1}$ to $\mathcal{D}$ is dissipative in $L^{2}(H, \rho v)$.

We remark that $\mathcal{D}$ is dense in $L^{2}(H, \rho \nu)$ since it contains $\mathcal{F C}_{b}^{\infty}(H)$ which is dense by the extension of Lemma 2.2 to $L^{2}(H, \rho \nu)$. Moreover $\left(\lambda I-K_{1}\right)(\mathcal{D})$ is dense for $\lambda>\omega_{0}$, since it contains $\mathcal{F C}_{b}^{\infty}(H)$. Indeed, if $f \in \mathcal{F} \mathcal{C}_{b}^{\infty}(H)$, then $u=$ $R\left(\lambda, K_{1}\right) f$ belongs to $\mathcal{D}$ and $\lambda u-K_{1} u=f$.

Let us denote by $\widetilde{K}_{1}: D\left(\widetilde{K}_{1}\right) \mapsto L^{2}(H, \rho \nu)$ the closure of $K_{1}: \mathcal{D} \mapsto L^{2}(H, \rho \nu)$. By the Lumer-Phillips theorem, $\widetilde{K}_{1}$ generates a strongly continuous contraction semigroup in $L^{2}(H, \rho v)$, and $\mathcal{D}$ is a core for $\widetilde{K}_{1}$. So, for every $\varphi \in D\left(\widetilde{K}_{1}\right)$ there is a sequence of functions $\varphi_{n} \in \mathcal{D}$ such that $\varphi_{n} \rightarrow \varphi$ and $K_{1} \varphi_{n} \rightarrow \widetilde{K}_{1} \varphi$ in $L^{2}(H, \rho \nu)$. For every $n$ we have

$$
\int_{H} K_{1} \varphi_{n} \rho d v=\int_{H} \varphi_{n} K_{1}^{*} \rho d v=0
$$

and letting $n \rightarrow \infty$ we obtain $\int_{H} \widetilde{K}_{1} \varphi \rho d \nu=0$. This proves the last statement.
5. Kolmogorov equations of stochastic reaction-diffusion equations. Let $H=L^{2}((0,1), d \xi)$, and let $A$ be the realization of the second order derivative with Dirichlet boundary condition, that is, $D(A)=W^{2,2}((0, \pi), d \xi) \cap$ $W_{0}^{1,2}((0, \pi), d \xi), A x=x^{\prime \prime}$.

We consider the Gaussian measure $\mu$ in $H$ with mean 0 and covariance $Q:=-\frac{1}{2} A^{-1}$. A canonical orthonormal basis of $H$ consists of the functions $e_{k}(\xi):=\sqrt{2} \sin (k \pi \xi), k \in \mathbb{N}$, that are eigenfunctions of $Q$ with eigenvalues $\lambda_{k}:=1 /\left(2 k^{2} \pi^{2}\right)$.

Let $\Phi: \mathbb{R} \mapsto \mathbb{R}$ be any convex lowerly bounded function, with (at most) polynomial growth at infinity, say

$$
\begin{equation*}
|\Phi(t)| \leq C\left(1+|t|^{p_{1}}\right), \quad t \in \mathbb{R} \tag{5.1}
\end{equation*}
$$

for some $C>0, p_{1} \geq 2$. We set

$$
U(x)= \begin{cases}\int_{0}^{1} \Phi(x(\xi)) d \xi, & x \in L^{p_{1}}(0,1)  \tag{5.2}\\ +\infty, & x \notin L^{p_{1}}(0,1)\end{cases}
$$

Section 5.1 is devoted to check that $U$ satisfies Hypotheses 2.1 and 2.8 , so that we can apply Theorem 3.9 to obtain regularity results for the solution $u$ to (1.1). Then in Section 5.2 we show that under an additional assumption $u$ fulfills (3.33) too.
5.1. Checking Hypotheses 2.1 and 2.8. We first note that $U$ is finite $\mu$-a.e., thanks to the next lemma. Its statement should be well known; however, we write down a simple proof for the reader's convenience.

Lemma 5.1. For every $p \geq 2$ we have

$$
\begin{equation*}
\int_{H} \int_{0}^{1}|x(\xi)|^{p} d \xi d \mu<\infty \tag{5.3}
\end{equation*}
$$

and hence $\mu\left(L^{p}(0,1)\right)=1$. Moreover, $x \mapsto\|x\|_{L^{p}(0,1)} \in L^{q}(H, \mu)$ for every $q \geq 1$.

Proof. Let $P_{n}$ be the orthogonal projection on the subspace spanned by $e_{1}, \ldots, e_{n}$. For every $\xi \in(0,1)$ and $m<n \in \mathbb{N}$, the function $x \mapsto P_{n} x(\xi)-$


$$
\begin{aligned}
\int_{H}\left|P_{n} x(\xi)-P_{m} x(\xi)\right|^{p} d \mu & =\int_{\mathbb{R}}|\eta|^{p} N_{0, \sum_{k=m+1}^{n} \lambda_{k} e_{k}(\xi)^{2}}(d \eta) \\
& =c_{p}\left(\sum_{k=m+1}^{n} \lambda_{k} e_{k}(\xi)^{2}\right)^{p / 2} \\
& \leq \tilde{c}_{p}\left(\sum_{k=m+1}^{n} \lambda_{k}\right)^{p / 2}
\end{aligned}
$$

with $\tilde{c}_{p}=2^{p / 2} c_{p}$, so that

$$
\begin{aligned}
\int_{H} \int_{0}^{1}\left|P_{n} x(\xi)-P_{m} x(\xi)\right|^{p} d \xi d \mu & =\int_{0}^{1} \int_{H}\left|P_{n} x(\xi)-P_{m} x(\xi)\right|^{p} d \mu d \xi \\
& \leq \tilde{c}_{p}\left(\sum_{k=m+1}^{n} \lambda_{k}\right)^{p / 2}
\end{aligned}
$$

This implies that the sequence $(x, \xi) \mapsto P_{n} x(\xi)$ converges in $L^{p}(H \times(0,1), \mu \times$ $d \xi$ ) to a limit function $u$ that belongs to $L^{p}(H \times(0,1), \mu \times d \xi)$ for every $p$. Let us show that $u(x, \xi)=x(\xi)$ taking $p=2$ : indeed, $\int_{0}^{1}\left|P_{n} x(\xi)-x(\xi)\right|^{2} d \xi$ vanishes for every $x \in H$ as $n \rightarrow \infty$, and it is bounded by $\|x\|^{2}$ which belongs to $L^{1}(H, \mu)$, so that by the dominated convergence theorem, $\int_{H} \int_{0}^{1}\left|P_{n} x(\xi)-x(\xi)\right|^{2} d \xi d \mu$ vanishes as $n \rightarrow \infty$. Then $u(x, \xi)=x(\xi)$, and (5.3) follows. It implies that $\mu\left(L^{p}(H, \mu)\right)=1$ for every $p \geq 2$ and that $x \mapsto\|x\|_{L^{p}(0,1)} \in L^{p}(H, \mu)$. For $q>p$ and $x \in L^{q}(0,1)$, the Hölder inequality yields $\|x\|_{L^{p}(0,1)} \leq\|x\|_{L^{q}(0,1)}$ so that $x \mapsto\|x\|_{L^{p}(0,1)} \in L^{q}(H, \mu)$.

The function $U$ defined by (5.2) is convex and bounded from below because $\Phi$ is. Using the Fatou lemma, it is easily seen to be lowerly semicontinuous. By assumption (5.1) and Lemma 5.1, $U \in L^{p}(H, \mu)$ for every $p \geq 1$, and the measures
$\mu$ and $\nu=e^{-2 U} \mu / \int_{H} e^{-2 U} d \mu$ are equivalent. For $U$ belong to some Sobolev space it is sufficient that also $\Phi^{\prime}$ has at most polynomial growth, as the next proposition shows.

Proposition 5.2. Let $\Phi: \mathbb{R} \mapsto \mathbb{R}$ be any $C^{1}$ convex lowerly bounded function such that

$$
\begin{equation*}
\left|\Phi^{\prime}(t)\right| \leq C\left(1+|t|^{p_{2}}\right), \quad t \in \mathbb{R} \tag{5.4}
\end{equation*}
$$

for some $C>0, p_{2} \geq 1$. Then the function $U$ defined in (5.2) belongs to $W_{0}^{1, p}(H, \mu)$ for every $p \geq 1$, and $D U(x)=\Phi^{\prime} \circ x$ for a.e. $x \in H$ [namely, for each $\left.x \in L^{2 p_{2}}(0,1)\right]$.

Proof. By (5.4), $\Phi$ satisfies (5.1) with $p_{1}=p_{2}+1$, so that $U \in L^{p}(H, \mu)$ for every $p$ by Lemma 5.1. To prove that $U \in W_{0}^{1, p}(H, \mu)$ we shall approach $U$ by its Moreau-Yosida approximations $U_{\alpha}$ defined in (2.9). Each $U_{\alpha}$ is continuously differentiable and $D U_{\alpha}$ is Lipschitz continuous, hence $U_{\alpha} \in W_{0}^{1, p}(H, \mu)$ for every $p$. This can be easily proved arguing as in the case $p=2$ of [8], Proposition 10.11.

Since $U_{\alpha}(x)$ converges monotonically to $U(x)$ at each $x$ such that $U(x)<\infty$, by Lemma $5.1 U_{\alpha}$ converges to $U$, $\mu$-a.e. Since

$$
\begin{aligned}
\inf U & \leq U_{\alpha}(x) \leq U(x) \\
& \leq C\left(1+\int_{0}^{1}|x(\xi)|^{p_{1}} d \xi\right) \\
& \leq C\left(1+\left(\int_{0}^{1}|x(\xi)|^{p_{1} p} d \xi\right)^{1 / p}\right)
\end{aligned}
$$

by Lemma 5.1 and the dominated convergence theorem, $U_{\alpha} \rightarrow U$ in $L^{p}(H, \mu)$.
Let $x \in L^{2 p_{2}}(0,1)$. Then the subdifferential $\partial U(x)$ is not empty. Indeed, since $\Phi$ is convex, for each $y \in H$ we have

$$
\begin{align*}
U(y)-U(x) & =\int_{0}^{\pi}[\Phi(x(\xi))-\Phi(y(\xi))] d \xi \\
& \geq \int_{0}^{\pi} \Phi^{\prime}(x(\xi))(x(\xi)-y(\xi)) d \xi \tag{5.5}
\end{align*}
$$

which implies that the function $\Phi^{\prime} \circ x \in H$ belongs to $\partial U(x)$. In fact, $\Phi^{\prime} \circ x \in$ $H$ is the unique element of $\partial U(x)$; see, for example, [2], Proposition 2.5. By Lemma 5.1, $x \mapsto\left\|\Phi^{\prime} \circ x\right\| \in L^{p}(H, \mu)$, and again by the dominated convergence theorem $\int_{H}\left\|D U_{\alpha}(x)-\Phi^{\prime} \circ x\right\|^{p} d \mu \rightarrow 0$ as $\alpha \rightarrow 0$, which shows that $U \in W_{0}^{1, p}(H, \mu)$ and $D U(x)=\Phi^{\prime} \circ x, \mu$-a.e.

If the assumptions of Proposition 5.2 hold, then $U$ satisfies Hypotheses 2.1 and 2.8, and consequently the results of Theorem 3.9 and of Propositions 4.7 and 4.8 hold.
5.2. Further estimates of $D u$. We are going to show that for every $\lambda>0$ and $f \in L^{2}(H, v)$, the solution of (1.1) satisfies estimate (3.33) as well, under reasonable additional assumptions on $\Phi$. We use the following preliminary result.

Proposition 5.3. Let $g \in C^{2}(\mathbb{R})$ be such that

$$
\begin{equation*}
\left|g^{\prime \prime}(t)\right| \leq C\left(1+|t|^{m}\right), \quad t \in \mathbb{R} \tag{5.6}
\end{equation*}
$$

for some $C>0, m \geq 1$. Then the function $F(x):=g \circ x$ belongs to $W_{1 / 2}^{1, q}(H, \mu ; H)$ for all $q>1$. If in addition $g_{\alpha}: \mathbb{R} \mapsto \mathbb{R}$ are $C^{2}$ functions fulfilling (5.6) with constant $C$ independent of $\alpha>0$ and $g_{\alpha}, g_{\alpha}^{\prime}$ pointwise converge to $g, g^{\prime}$, respectively, as $\alpha \rightarrow 0^{+}$, then $F_{\alpha}(x):=g_{\alpha} \circ x$ converges to $F$ in $W_{1 / 2}^{1, q}(H, \mu ; H)$ as $\alpha \rightarrow 0^{+}$ for all $q>1$.

Proof. As first step we show that for each $x \in L^{2 m}(0,1)$ (hence, $\mu$-a.e.), $F$ is differentiable in any direction $h \in Q^{1 / 2}(H)=H_{0}^{1}(0,1)$ and that $\frac{\partial F(x)}{\partial h}=g^{\prime} \circ x \cdot h$. We have in fact for all $h \in H_{0}^{1}(0,1), \xi \in(0,1)$ and all $0<|t| \leq 1$,

$$
\begin{aligned}
& \left|\frac{g(x+t h)(\xi)-g(x(\xi))}{t}-g^{\prime}(x(\xi)) h(\xi)\right| \\
& \quad=\left|\int_{0}^{1}\left[g^{\prime}(x(\xi)+t \sigma h(\xi))-g^{\prime}(x(\xi))\right] h(\xi) d \sigma\right| \\
& \quad=\left|\int_{0}^{1} \int_{0}^{1} g^{\prime \prime}(x(\xi)+t \sigma \eta h(\xi)) t \sigma h(\xi)^{2} d \eta d \sigma\right| \\
& \quad \leq t\|h\|_{\infty}^{2} C\left(1+2^{m-1} \mid\left(|x(\xi)|^{m}+\|h\|_{\infty}^{m}\right)\right) .
\end{aligned}
$$

Now, taking the square and integrating over $(0,1)$, yields

$$
\left\|\frac{F(x+t h)-F(x)}{t}-g^{\prime} \circ x \cdot h\right\|_{H} \leq t C(h)\left(1+\|x\|_{L^{2 m}}^{m}\right) .
$$

This implies that for each $x \in L^{2 m}(0,1), F$ is differentiable at $x$ in any direction $h \in H_{0}^{1}(0,1)$ and that

$$
\frac{\partial F(x)}{\partial h}=g^{\prime} \circ x \cdot h
$$

Let us notice that $F, \partial F / \partial h$ belong to $L^{q}(H, \mu ; H)$ for every $q \geq 1$. Indeed, (5.6) implies that $|g(t)| \leq M\left(1+|t|^{m+2}\right),\left|g^{\prime}(t)\right| \leq M\left(1+|t|^{m+1}\right)$ for every $t \in \mathbb{R}$ and for some $M>0$, so that $|F(x(\xi))| \leq M\left(1+|x(\xi)|^{m+2}\right),|\partial F(x) / \partial h(\xi)| \leq M(1+$ $\left.|x(\xi)|^{m+1}\right)\|h\|_{\infty}$ and then

$$
\begin{aligned}
\|F(x)\|_{H}^{2} & \leq \int_{0}^{1} M^{2}\left(1+|x(\xi)|^{m+2}\right)^{2} d \xi \\
\left\|\frac{\partial F(x)}{\partial h}(x)\right\|_{H}^{2} & \leq\|h\|_{\infty}^{2} \int_{0}^{1} M^{2}\left(1+|x(\xi)|^{m+1}\right)^{2} d \xi
\end{aligned}
$$

and the right-hand sides belong to $L^{q}(H, \mu)$ for every $q$. It follows from [3], Section 5.2 , that $F$ belongs to $G^{q, 1}(H, \mu ; H)$ [i.e., $F$ belongs to $L^{q}(H, \mu ; H)$, it is weakly differentiable in all directions of the Cameron-Martin space $H_{0}^{1}(0,1)$ and any weak derivative $\frac{\partial F(x)}{\partial h}$ with $h \in H_{0}^{1}(0,1)$ can be expressed as $\Psi(x) h$, where $\Psi \in L^{q}\left(H, \mu ; \mathcal{L}\left(H_{0}^{1}(0,1), H\right)\right)$ is such that $\left.\partial F(x) / \partial h=\Psi(x)(h)\right]$. To show that $F \in W_{1 / 2}^{1, q}(H, \mu ; H)$ we have still to check that ([3], Proposition 5.4.6, Corollary 5.4.7)

$$
\int_{H}\left(\sum_{h, k \in \mathbb{N}} \lambda_{h} \lambda_{k}\left\langle\partial F(x) / \partial e_{h},\left.e_{k}\right|^{2}\right)^{q / 2} d \mu<\infty\right.
$$

This is because a canonical orthonormal basis of $H_{0}^{1}(0,1)$ is just the set $\left\{\sqrt{\lambda_{k}} e_{k}: k \in \mathbb{N}\right\}$. Recalling that $\left\|e_{k}\right\|_{\infty}=\sqrt{2}$ for every $k$, we get

$$
\begin{aligned}
\left|\left\langle\partial F(x) / \partial e_{h}, e_{k}\right\rangle\right| & =\left|\int_{0}^{1} g^{\prime}(x(\xi)) e_{h}(\xi) e_{k}(\xi) d \xi\right| \\
& \leq 2 M \int_{0}^{1}\left(1+|x(\xi)|^{m+1}\right) d \xi \\
& =2 M\left(1+\|x\|_{L^{m+1}}^{m+1}\right)
\end{aligned}
$$

for each $h, k \in \mathbb{N}$, which implies

$$
\int_{H}\left(\sum_{h, k \in \mathbb{N}} \lambda_{h} \lambda_{k}\left\langle\partial F(x) / \partial e_{h}, e_{k}\right\rangle^{2}\right)^{q / 2} d \mu \leq 2 M \int_{H}(\operatorname{Tr} Q)^{q}\|x\|_{L^{m+1}}^{q(m+1)} d \mu<\infty
$$

so that $F \in W_{1 / 2}^{1, q}(H, \mu ; H)$.
Now we can show that $F_{\alpha} \rightarrow F$ as $\alpha \rightarrow 0$. In fact, since (5.6) is fulfilled with constant independent of $\alpha$, there is $M_{1}>0$ independent of $\alpha$ such that

$$
\left|g_{\alpha}(t)\right| \leq M_{1}\left(1+|t|^{m+2}\right), \quad\left|g_{\alpha}^{\prime}(t)\right| \leq M_{1}\left(1+|t|^{m+1}\right), \quad t \in \mathbb{R}
$$

Concerning the convergence of $g_{\alpha} \circ x$ to $g \circ x$ in $L^{q}(H, \mu ; H)$ we have

$$
\begin{aligned}
\int_{H}\left\|g_{\alpha} \circ x-g \circ x\right\|_{H}^{q} d \mu & =\int_{H}\left(\int_{0}^{1}\left|g_{\alpha}(x(\xi))-g(x(\xi))\right|^{2} d \xi\right)^{q / 2} d \mu \\
& \leq \int_{H} \int_{0}^{1}\left|g_{\alpha}(x(\xi))-g(x(\xi))\right|^{q} d \xi d \mu
\end{aligned}
$$

and the last integral goes to 0 as $\alpha \rightarrow 0$ by the dominated convergence theorem. Therefore $F_{\alpha}(x)=g_{\alpha} \circ x$ converges to $F$ in $L^{q}(H, \mu ; H)$. Concerning the con-
vergence in $W_{1 / 2}^{1, q}(H, \mu ; H)$ we have

$$
\begin{aligned}
& \int_{H}\left(\sum_{h, k \in \mathbb{N}} \lambda_{h} \lambda_{k}\left\langle\partial\left(g_{\alpha} \circ x\right) / \partial e_{h}-\partial(g \circ x) / \partial e_{h}, e_{k}\right\rangle^{2}\right)^{q / 2} d \mu \\
& \quad=\int_{H}\left(\sum_{h, k \in \mathbb{N}} \lambda_{h} \lambda_{k}\left(\int_{0}^{1}\left(g_{\alpha}^{\prime}(x(\xi))-g^{\prime}(x(\xi))\right) e_{h}(\xi) e_{k}(\xi) d \xi\right)^{2}\right)^{q / 2} d \mu \\
& \quad \leq C_{q} \int_{H}\left(\sum_{h, k \in \mathbb{N}} \lambda_{h} \lambda_{k} \int_{0}^{1}\left|g_{\alpha}^{\prime}(x(\xi))-g^{\prime}(x(\xi))\right|^{2} d \xi\right)^{q / 2} d \mu \\
& \quad \leq C_{q}(\operatorname{Tr} Q)^{q} \int_{H} \int_{0}^{1}\left|g_{\alpha}^{\prime}(x(\xi))-g^{\prime}(x(\xi))\right|^{q} d \xi d \mu
\end{aligned}
$$

and the last integral vanishes as $\alpha \rightarrow 0$ again by the dominated convergence theorem.

We shall use Proposition 5.3 to prove that the Moreau-Yosida approximations $U_{\alpha}$ converge to $U$ in $W_{1 / 2}^{2, q}(H, \mu)$ for every $q$ [for the moment, we only know convergence in $\left.W^{1, q}(H, \mu)\right]$.

Proposition 5.4. Let $\Phi: \mathbb{R} \mapsto \mathbb{R}$ be any $C^{3}$ convex lowerly bounded function such that

$$
\begin{equation*}
\left|\Phi^{\prime \prime \prime}(t)\right| \leq C\left(1+|t|^{m}\right), \quad t \in \mathbb{R} \tag{5.7}
\end{equation*}
$$

for some $C, m>0$. Then $U \in W_{1 / 2}^{2, q}(H, \mu)$ for all $q>1$, and we have

$$
\lim _{\alpha \rightarrow 0} U_{\alpha}=U \quad \text { in } W_{1 / 2}^{2, q}(H, \mu) \forall q>1
$$

Proof. Let us apply Proposition 5.3 to $F(x)=D U(x)=g \circ x$ with $g=\Phi^{\prime}$. Since $g^{\prime \prime}$ has polynomial growth, $F \in W_{1 / 2}^{2, q}(H, \mu ; H)$ for all $q$, so that $U \in$ $W_{1 / 2}^{2, q}(H, \mu)$ for all $q$. Moreover $D U_{\alpha}(x)=D_{0} U\left(y_{\alpha}\right)$, where $y_{\alpha}$ is the solution of

$$
y_{\alpha}+\alpha D_{0} U\left(y_{\alpha}\right)=x
$$

that is

$$
y_{\alpha}+\alpha \Phi^{\prime}\left(y_{\alpha}\right)=x
$$

Therefore

$$
y_{\alpha}(\xi)=\left(I+\alpha \Phi^{\prime}\right)^{-1}(x(\xi)), \quad 0<\xi<1
$$

and so

$$
D U_{\alpha}(x)=\Phi^{\prime} \circ\left(I+\alpha \Phi^{\prime}\right)^{-1} \circ x
$$

Setting $g_{\alpha}(t)=\Phi^{\prime} \circ\left(I+\alpha \Phi^{\prime}\right)^{-1}(t)$, we see that $g_{\alpha}$ converges pointwise to $g=\Phi^{\prime}$, and

$$
g_{\alpha}^{\prime}=\frac{\Phi^{\prime \prime} \circ\left(I+\alpha \Phi^{\prime}\right)^{-1}}{\left(1+\alpha \Phi^{\prime \prime} \circ\left(I+\alpha \Phi^{\prime}\right)^{-1}\right)}
$$

converges pointwise to $g^{\prime}=\Phi^{\prime \prime}$.
Moreover we notice that there exists $M>0$, independent of $\alpha \in(0,1)$ such that $\left|\left(I+\alpha \Phi^{\prime}\right)^{-1}(t)\right| \leq M+|t|$ for all $t \in \mathbb{R}$. (5.7) implies that $\Phi^{\prime}$ and $\Phi^{\prime \prime}$ have polynomial growth as well; in particular $\left|\Phi^{\prime}(t)\right| \leq c_{1}\left(1+|t|^{m+2}\right)$, so that $\left|g_{\alpha}(t)\right| \leq c_{1}\left(1+(M+|t|)^{m+2}\right)$. A similar estimate with $m+1$ instead of $m+2$ holds also for $\left|g_{\alpha}^{\prime}(t)\right|$. By the second part of Proposition 5.3, $D U_{\alpha}$ converges to $D U$ in $W_{1 / 2}^{1, q}(H, \mu ; H)$ as $\alpha \rightarrow 0$, thereby $U_{\alpha}$ converges to $U$ in $W_{1 / 2}^{2, q}(H, \mu)$.

As a final step, we can show that the solution to (1.1) satisfies (3.33) under the assumptions of Proposition 5.4.

PROPOSITION 5.5. Let $U$ be defined by (5.2) with $\Phi: \mathbb{R} \mapsto \mathbb{R}$ convex, bounded from below, of class $C^{3}$ and satisfying (5.7). Then for every $\lambda>0$ and $f \in L^{2}(H, v)$ the weak solution $u$ of (1.1) satisfies (3.33).

Proof. It is sufficient to prove the statement for $f \in C_{b}(H)$, which is dense in $L^{2}(H, v)$. By Lemma 3.10 there is a sequence $\left(\alpha_{n}\right) \rightarrow 0$ such that $u_{\alpha_{n}} \rightarrow u$ in $W^{1,2}(H, v)$. Then $D u_{\alpha_{n}} \rightarrow D u$ in $L^{2}(H, v ; H)$ so that (up to a subsequence) $D u_{\alpha_{n}}(x) \rightarrow D u(x)$ for almost all $x$. By Proposition 5.4, $U_{\alpha_{n}}$ converges to $U$ in $W_{1 / 2}^{2,2}(H, \mu)$, thereby for all fixed $h, k \in \mathbb{N}$ we have $D_{h k} U_{\alpha_{n}} \rightarrow D_{h k} U$ in $L^{2}(H, \mu)$. Let us fix $N \in \mathbb{N}$. Possibly choosing a further subsequence, we have $D_{h k} U_{\alpha_{n}} \rightarrow$ $D_{h k} U$ pointwise a.e. for all $h, k \leq N$. Therefore for $\mu$-a.e. $x \in H$ we have

$$
\begin{gathered}
\lim _{n \rightarrow \infty} \sum_{h, k=1}^{N} D_{h k} U_{\alpha_{n}}(x) D_{h} u_{\alpha_{n}}(x) D_{k} u_{\alpha_{n}}(x) e^{-2 U_{\alpha_{n}}(x)} \\
\quad=\sum_{h, k=1}^{N} D_{h k} U(x) D_{h} u(x) D_{k}(x) e^{-2 U(x)}
\end{gathered}
$$

and by Fatou's lemma,

$$
\begin{aligned}
& \int_{H} \sum_{h, k=1}^{N} D_{h k} U(x) D_{h} u(x) D_{k} u(x) d v \\
& \quad=\int_{H} \sum_{h, k=1}^{N} D_{h k} U(x) D_{h} u(x) D_{k} u(x) e^{-2 U(x)} d \mu
\end{aligned}
$$

$$
\begin{aligned}
& \leq \liminf _{n \rightarrow \infty} \int_{H} \sum_{h, k=1}^{N} D_{h k} U_{\alpha_{n}}(x) D_{h} u_{\alpha_{n}}(x) D_{k} u_{\alpha_{n}}(x) e^{-2 U_{\alpha_{n}}(x)} d \mu \\
& \leq 4 \liminf _{n \rightarrow \infty} \int_{H} f^{2} e^{-2 U_{\alpha_{n}}} d \mu=4 \int_{H} f^{2} d \nu
\end{aligned}
$$

Now by Theorem 3.9 we know that $x \mapsto\|D u(x)\|_{H_{0}^{1}(0,1)}=\left\|Q^{-1 / 2} D u(x)\right\|_{H} / \sqrt{2}$ $\in L^{2}(H, \mu)$, therefore for almost any $x \in H, D u(x) \in H_{0}^{1}(0,1)$, whereas by Proposition 5.4 it follows that $x \mapsto \sum_{h, k=1}^{\infty} \lambda_{h} \lambda_{k}\left(D_{h k} U(x)\right)^{2}$ belongs to $L^{1}(H, \mu)$, that is $x \mapsto\left\|D^{2} U(x)\right\|_{\mathcal{L}_{2}\left(H_{0}^{1}(0,1)\right)} \in L^{2}(H, \mu)$. Therefore for almost $x \in H$, $D^{2} U(x) \in \mathcal{L}_{2}\left(H_{0}^{1}(0,1)\right)$. It follows that for almost any $x \in H$ the sequence $\sum_{h, k=1}^{N} D_{h k} U(x) D_{k} u(x) D_{k} u(x)$ converges to $\sum_{h, k=1}^{\infty} D_{h k} U(x) D_{k} u(x) D_{k} u(x)$. Using once again Fatou's lemma we can conclude that

$$
\begin{aligned}
& \int_{H} \sum_{h, k=1}^{\infty} D_{h k} U(x) D_{h} u(x) D_{k} u(x) d v \\
& \quad=\int_{H} \lim _{N \rightarrow \infty} \sum_{h, k=1}^{N} D_{h k} U(x) D_{h} u(x) D_{k} u(x) d v \\
& \quad \leq \liminf _{N \rightarrow \infty} \int_{H} \sum_{h, k=1}^{N} D_{h k} U(x) D_{h} u(x) D_{k} u(x) d v \leq 4 \int_{H} f^{2} d v
\end{aligned}
$$

Then we can apply all the results of Sections 3 and 4. In particular, we have the following theorem.

THEOREM 5.6. Let $\Phi: \mathbb{R} \mapsto \mathbb{R}$ be any convex $C^{1}$ lowerly bounded function satisfying (5.4), and let $U$ be defined by (5.2). Then for every $\lambda>0$ and $f \in L^{2}(H, v)$ the weak solution $u$ to (1.1) belongs to $W^{2,2}(H, v) \cap W_{-1 / 2}^{1,2}(H, v)$, and it satisfies (3.27), (3.28). If in addition $\Phi$ is $C^{3}$ and satisfies (5.7), then $u$ satisfies (3.33) as well.

With our choice of $U$, the stochastic differential equation (1.2) in $H$ reads as

$$
\begin{equation*}
d X=\left(A X-\Phi^{\prime}(X)\right) d t+d W(t), \quad X(0)=x \tag{5.8}
\end{equation*}
$$

and hence it is a reaction-diffusion SPDE, whose Kolmogorov operator is just $\mathcal{K}$. As in Section 3.2.4, $W(t)$ is any $H$-valued cylindrical Wiener process defined in a probability space $(\Omega, \mathcal{F}, \mathbb{P})$. The connection between (5.8) and (1.1) is stated in the next proposition. The definition of mild solution to (5.8) is the same as in the case of Lipschitz continuous $D U$.

Proposition 5.7. Let $\Phi: \mathbb{R} \mapsto \mathbb{R}$ be a convex lowerly bounded function satisfying (5.4) for some $p_{2} \geq 1$. Then for every $x \in L^{2 p_{2}}(0,1)$ (hence, for $\mu$-a.e. $x \in H$ ) problem (5.8) has a unique mild solution $X$. For every $f \in C_{b}(H)$ we have

$$
\begin{equation*}
u(x)=\int_{0}^{\infty} e^{-\lambda t} \mathbb{E}(f(X(t, x))) d t \tag{5.9}
\end{equation*}
$$

$\mu$-a.e. $x \in H$, where $u$ is the weak solution to (1.1).
Proof. Existence of a unique mild solution to (5.8) follows from [14], Theorem 5.5.8, that deals with Cauchy problems such as $d X=(A X+F(X)) d t+$ $d W(t), X(0)=x$. In our case, $F(x)=-D U(x)=-\Phi^{\prime}(x)$ satisfies the assumptions of [14], Theorem 5.5.8, with $K=L^{2 p_{2}}(0,1)$. In particular, Hypothesis 5.5 is satisfied, since in [7], Proposition 4.3, it is proved that $(t, \xi) \mapsto$ $\int_{0}^{t} e^{(t-s) A} d W(s)(\xi)$ is a.s. continuous.

The mild solution is obtained as the limit of mild solutions to approximating problems,

$$
d X_{\alpha}=\left(A X_{\alpha}-D U_{\alpha}(X)\right) d t+d W(t), \quad X(0)=x
$$

as $\alpha \rightarrow 0$, where $D U_{\alpha}$ are the Yosida approximations of $D U$, and for each $T>0$ we have $\lim _{\alpha \rightarrow 0} \sup _{0 \leq t \leq T}\left\|X_{\alpha}(t)-X(t)\right\|=0, \mathbb{P}$-a.e. By Proposition 3.8, for every $\lambda>0$,

$$
\begin{equation*}
R\left(\lambda, K_{\alpha}\right) f=\int_{0}^{\infty} e^{-\lambda t} \mathbb{E}\left(f\left(X_{\alpha}(t, \cdot)\right)\right) d t \tag{5.10}
\end{equation*}
$$

We recall that $R\left(\lambda, K_{\alpha}\right) f=u_{\alpha}$ is the weak solution to (3.29), and that a sequence $u_{\alpha_{n}}$ with $\alpha_{n} \rightarrow 0$ converges to $u$ in $L^{2}(H, \mu)$ as $n \rightarrow \infty$, by Lemma 3.10. Moreover, $\int_{0}^{\infty} e^{-\lambda t} \mathbb{E}\left(f\left(X_{\alpha_{n}}(t, \cdot)\right)\right) d t$ goes to $\int_{0}^{\infty} e^{-\lambda t} \mathbb{E}(f(X(t, x)) d t$ pointwise $\mu$-a.e. and also in $L^{2}(H, \mu)$, by the dominated convergence theorem. Taking $\alpha=\alpha_{n}$ in (5.10) and letting $n \rightarrow \infty$ formula (5.9) follows.

Concerning perturbed equations,

$$
\begin{equation*}
d X=\left(A X-\Phi^{\prime}(X)+B(X)\right) d t+d W(t) \tag{5.11}
\end{equation*}
$$

we do not know about existence of invariant measures except in the case of bounded perturbations of Ornstein-Uhlenbeck equations. See [14], Chapter 8. If $B$ is a bounded Borel function, Proposition 4.8 yields that the corresponding Kolmogorov semigroup $e^{t K_{1}}$ has an invariant measure $\nu$. The verification of formula (5.9) where now $X(t, x)$ is the mild solution to (5.11) and $u=R\left(\lambda, K_{1}\right)$ is not obvious. In fact, even existence of a mild solution is not obvious. It could be done through the Girsanov transform, but the argument is quite delicate and we hope to be able to treat the subject in a future paper.
6. Kolmogorov equations of stochastic Cahn-Hilliard-type problems. In Section 5 we have seen that the superposition $x \mapsto \Phi^{\prime} \circ x$ may be seen as the gradient of a suitable function $U$ in the space $L^{2}(0,1)$. This is no longer true for
operators of the type $x \mapsto \frac{d}{d \xi}\left(\Phi^{\prime} \circ x\right)$ or $x \mapsto \frac{d^{2}}{d \xi^{2}}\left(\Phi^{\prime} \circ x\right)$. However they may be still interpreted as gradients, with suitable choices of the space $H$.

Here we set $V:=\left\{x \in H^{1}(0,1): \int_{0}^{1} x(\xi) d \xi=0\right\}$, with scalar product $\langle x, y\rangle_{V}=$ $\int_{0}^{1} x^{\prime}(\xi) y^{\prime}(\xi) d \xi$, and we choose $H$ to be the dual space of $V$, endowed with the dual norm. We consider the spaces $\tilde{L}^{p}(0,1):=\left\{x \in L^{p}(0,1): \int_{0}^{1} x(\xi) d \xi=0\right\}$ as subspaces of $H$, identifying any $x \in L^{p}(0,1)$ with zero mean value with the element $y \mapsto \int_{0}^{1} x(\xi) y(\xi) d \xi$ of $H$.

The standard extension $B$ of the negative second order derivative on $V$ with values in $H$ is defined by

$$
B x(y)=\int_{0}^{1} x^{\prime}(\xi) y^{\prime}(\xi) d \xi, \quad y \in V
$$

If $x \in V \cap H^{2}(0,1)$ and $x^{\prime}(0)=x^{\prime}(1)=0$, then $B x(y)=-\int_{0}^{1} x^{\prime \prime}(\xi) y(\xi) d \xi$ so that, with the above identification, $B$ is an extension of (minus) the second order derivative with Neumann boundary condition. The operator $B$ is an isometry between $V$ and $H$, since $\|B x\|_{H}=\sup _{y \neq 0}\langle x, y\rangle_{V} /\|y\|_{V}=\|x\|_{V}$. Moreover, if $z \in \widetilde{L}^{2}(0,1)$ and $x \in V$, then $\langle z, B x\rangle_{H}=\langle z, x\rangle_{L^{2}(0,1)}$.

Let $e_{k}(\xi):=\sqrt{2} \cos (k \pi \xi)$. Then $\left\{e_{k}: k \in \mathbb{N}\right\}$ is an orthonormal basis of $\widetilde{L}^{2}(0,1), B e_{k}=k^{2} \pi^{2} e_{k}$, and setting $f_{k}=k \pi e_{k}$, the set $\left\{f_{k}: k \in \mathbb{N}\right\}$ is an orthonormal basis of $H$. We recall that $P_{n}$ is the orthogonal projection on the subspace spanned by the first $n$ elements of the basis,

$$
P_{n} x=\sum_{k=1}^{n}\left\langle x, f_{k}\right\rangle_{H} f_{k}
$$

REMARK 6.1. Note that the restriction of $P_{n}$ to $\widetilde{L}^{2}(0,1)$ is the orthogonal projection in $\widetilde{L}^{2}(0,1)$ on the subspace spanned by $e_{1}, \ldots, e_{n}$. Indeed, for every $x \in \widetilde{L}^{2}(0,1)$ and $k \in \mathbb{N}$ we have

$$
\left\langle x, f_{k}\right\rangle_{H} f_{k}=\left\langle x, B^{-1} f_{k}\right\rangle_{L^{2}} f_{k}=\left\langle x, \frac{e_{k}}{k \pi}\right\rangle_{L^{2}} k \pi e_{k}=\left\langle x, e_{k}\right\rangle_{L^{2}} e_{k} .
$$

Here we set $A=-B^{2}$ and, as usual, we denote by $\mu$ the Gaussian measure on $H$ with zero mean and covariance $Q=-A^{-1} / 2$. Note that the eigenvalues of $Q$ are now $\lambda_{k}:=1 / 2 \pi^{4} k^{4}$, and $B=\sqrt{2} Q^{1 / 2}$.

We consider a function $\Phi: \mathbb{R} \mapsto \mathbb{R}$ satisfying the following assumptions.
HYPOTHESIS 6.2. $\quad \Phi: \mathbb{R} \mapsto \mathbb{R}$ is a $C^{1}$ convex lowerly bounded function, satisfying (5.4) and

$$
\begin{equation*}
\lim _{r \rightarrow \pm \infty} \frac{\Phi(r)}{|r|}=+\infty \tag{6.1}
\end{equation*}
$$

Setting $p_{1}=p_{2}+1$, we define $U$ as in Section 5.1, by

$$
U(x)= \begin{cases}\int_{0}^{1} \Phi(x(\xi)) d \xi, & x \in \widetilde{L}^{p_{1}}(0,1)  \tag{6.2}\\ +\infty, & x \notin \widetilde{L}^{p_{1}}(0,1)\end{cases}
$$

$U$ is obviously convex and bounded from below, moreover by [2], Proposition 2.8, it is lower semicontinuous. To be more precise, in [2] the space $H$ is the dual space of $H_{0}^{1}(0,1)$, but the argument goes as well in our case. The subdifferential of $U$ is not empty at each $x \in \widetilde{L}^{1}(0,1)$ such that $\Phi^{\prime} \circ x \in V$ and it consists of the unique element $D_{0} U(x)=B\left(\Phi^{\prime} \circ x\right)$.

We shall see that $U \in W_{1 / 2}^{1,2}(H, \mu)$, while $U \notin W_{0}^{1,2}(H, \mu)$. For the proof, instead of approaching $U$ by its Moreau-Yosida approximations, we shall approach it by the sequence $U \circ P_{n}$; namely we set

$$
U_{n}(x)=\int_{0}^{1} \Phi\left(P_{n} x(\xi)\right) d \xi, \quad x \in H
$$

By (5.4), $\Phi$ satisfies (5.1), and we have $U(x) \leq C\left(1+\|x\|_{L^{p_{1}(0,1)}}^{p_{1}}\right), U_{n}(x) \leq$ $C\left(1+\left\|P_{n} x\right\|_{L^{p_{1}}(0,1)}^{p_{1}}\right)$. So, the starting point of our analysis is the study of the functions $x \mapsto\|x\|_{L^{p}(0,1)}, x \mapsto\left\|P_{n} x\right\|_{L^{p}(0,1)}$ for $p \geq 2$.

Proposition 6.3. For each $p \geq 1$ there is $C_{p}>0$ such that

$$
\begin{array}{r}
\int_{H} \int_{0}^{1}\left|P_{n} x(\xi)\right|^{p} d \xi d \mu \leq C_{p}\left(\sum_{k=1}^{n} \frac{1}{k^{2} \pi^{2}}\right)^{p / 2}, \quad n \in \mathbb{N},  \tag{6.3}\\
\int_{H} \int_{0}^{1}\left|P_{n} x(\xi)-P_{m} x(\xi)\right|^{p} d \xi d \mu \leq C_{p}\left(\sum_{k=m+1}^{n} \frac{1}{k^{2} \pi^{2}}\right)^{p / 2}, \\
m<n \in \mathbb{N} .
\end{array}
$$

Proof. First of all note that for every $x \in H, P_{n} x$ is a smooth function. Moreover for every $\xi \in(0,1)$ and $m<n \in \mathbb{N}$, the function $x \mapsto P_{n} x(\xi)-P_{m} x(\xi)$ is a


$$
\begin{aligned}
& \int_{H}\left|P_{n} x(\xi)-P_{m} x(\xi)\right|^{p} d \mu \\
& \quad=\int_{\mathbb{R}}|\eta|^{p} N_{0, \sum_{k=m+1}^{n}\left(1 /\left(\pi^{4} k^{4}\right)\right) f_{k}(\xi)^{2}}(d \eta) \\
& \quad=c_{p}\left(\sum_{k=m+1}^{n} \frac{1}{k^{2} \pi^{2}} e_{k}(\xi)^{2}\right)^{p / 2} \leq 2^{p / 2} c_{p}\left(\sum_{k=m+1}^{n} \frac{1}{k^{2} \pi^{2}}\right)^{p / 2}
\end{aligned}
$$

so that

$$
\begin{aligned}
& \int_{H} \int_{0}^{1}\left|P_{n} x(\xi)-P_{m} x(\xi)\right|^{p} d \xi d \mu \\
& \quad=\int_{0}^{1} \int_{H}\left|P_{n} x(\xi)-P_{m} x(\xi)\right|^{p} d \mu d \xi \leq 2^{p / 2} c_{p}\left(\sum_{k=m+1}^{n} \frac{1}{k^{2} \pi^{2}}\right)^{p / 2}
\end{aligned}
$$

that is, (6.4) holds. The proof of (6.3) is the same.
Proposition 6.3 has several consequences.
COROLLARY 6.4. $\quad \mu\left(\widetilde{L}^{p}(0,1)\right)=1$, and the sequence of functions $(x, \xi) \mapsto$ $P_{n} x(\xi)$ converges to $(x, \xi) \mapsto x(\xi)$ in $L^{p}(H \times(0,1), \mu \times d \xi)$, for every $p \geq 1$.

Proof. It is sufficient to prove that the statement holds for $p=2$. Indeed, estimate (6.4) implies that the sequence $(x, \xi) \mapsto P_{n} x(\xi)$ converges in $L^{p}(H \times$ $(0,1), \mu \times d \xi)$ for every $p$ to a limit function, that we identify with the function $(x, \xi) \mapsto x(\xi)$ taking $p=2$. Once we know that $\int_{H} \int_{0}^{1}|x(\xi)|^{p} d \xi d \mu<\infty$, then $\mu\left(\widetilde{L}^{p}(0,1)\right)$ is obviously 1 .

So, fix $p=2$. Since

$$
\begin{aligned}
& \int_{0}^{1}\left|P_{n} x(\xi)\right|^{2} d \xi \\
& \quad=\int_{0}^{1} \sum_{h, k=1}^{n}\left\langle x, f_{k}\right\rangle_{H}\left\langle x, f_{h}\right\rangle_{H} f_{k}(\xi) f_{h}(\xi) d \xi \\
& \quad=\int_{0}^{1} \sum_{k=1}^{n}\left\langle x, f_{k}\right\rangle_{H}^{2} f_{k}(\xi)^{2} d \xi
\end{aligned}
$$

then for every $x \in H$ the sequence $\int_{0}^{1}\left|P_{n} x(\xi)\right|^{2} d \xi$ is increasing, it converges to $\|x\|_{L^{2}}^{2}$ if $x \in \widetilde{L}^{2}(0,1)$, and to $+\infty$ if $x \notin \widetilde{L}^{2}(0,1)$ by Remark 6.1. By monotone convergence and (6.3) with $p=2$ the limit function belongs to $L^{1}(H, \mu)$, and this implies $\mu\left(\widetilde{L}^{2}(0,1)\right)=1$. Consequently, the function $(x, \xi) \mapsto x(\xi)$ is defined a.e. in $H \times(0,1)$. Moreover,

$$
\begin{aligned}
& \int_{\tilde{L}^{2}(0,1)} \int_{0}^{1}\left|P_{n} x(\xi)-x(\xi)\right|^{2} d \xi d \mu \\
& \quad= \\
& \quad \int_{\tilde{L}^{2}(0,1)} \lim _{m \rightarrow \infty} \int_{0}^{1}\left|P_{n} x(\xi)-P_{m} x(\xi)\right|^{2} d \xi d \mu \\
& \quad \leq \operatorname{limin}_{m \rightarrow \infty} \int_{\tilde{L}^{2}(0,1)} \int_{0}^{1}\left|P_{n} x(\xi)-P_{m} x(\xi)\right|^{2} d \xi d \mu
\end{aligned}
$$

For each $\varepsilon>0$ there is $n_{\varepsilon} \in \mathbb{N}$ such that for $n, m \geq n_{\varepsilon}$ we have $\int_{\tilde{L}^{2}(0,1)} \int_{0}^{1} \mid P_{n} x(\xi)-$ $\left.P_{m} x(\xi)\right|^{2} d \xi d \mu \leq \varepsilon$. Then for $n \geq n_{\varepsilon}$ we get $\int_{\tilde{L}^{2}(0,1)} \int_{0}^{1}\left|P_{n} x(\xi)-x(\xi)\right|^{2} d \xi d \mu \leq$ $\varepsilon$, and the statement follows.

Proposition 6.5. Under Hypothesis 6.2, $U \in W_{1 / 2}^{1, p}(H, \mu)$ and $\lim _{n \rightarrow \infty} U_{n}=U$ in $L^{p}(H, \mu)$, for every $p \geq 1$. Moreover, $D_{k} U(x)=$ $\int_{0}^{1} \Phi^{\prime}(x(\xi)) f_{k}(\xi) d \xi$ for a.e. $x \in H$.

Proof. As a first step, we remark that the sequence of functions $x \mapsto$ $\left\|P_{n} x\right\|_{L^{p}(0,1)}^{p}$ is bounded in $L^{s}(H, \mu)$ for every $s \geq 1$. Indeed, using the Hölder inequality we get

$$
\int_{0}^{1}\left|P_{n} x(\xi)\right|^{p} d \xi \leq\left(\int_{0}^{1}\left|P_{n} x(\xi)\right|^{p s} d \xi\right)^{1 / s}, \quad s \geq 1
$$

and the right-hand side belongs to $L^{s}(H, \mu)$ with norm independent of $n$, by estimate (6.3).

We already remarked that $\left|U_{n}(x)\right| \leq \int_{0}^{1} C\left(1+\left|P_{n} x(\xi)\right|\right)^{p_{1}} d \xi$ with $p_{1}=p_{2}+1$, so that $U_{n}$ is bounded in $L^{p}(H, \mu)$ by a constant independent of $n$, for every $p \geq 1$. Let us prove that $U_{n} \rightarrow U$ in $L^{p}(H, \mu)$. Using (5.4) and the Hölder inequality we get

$$
\begin{aligned}
\left|U_{n}(x)-U(x)\right|^{p} \leq & \left(\int_{0}^{1}\left|\Phi\left(P_{n} x(\xi)\right)-\Phi(x(\xi))\right| d \xi\right)^{p} \\
\leq & C^{p}\left(\int_{0}^{1}\left(1+|x(\xi)|+\left|P_{n} x(\xi)\right|\right)^{p_{2}}\left|P_{n} x(\xi)-x(\xi)\right| d \xi\right)^{p} \\
\leq & C^{p}\left(\int_{0}^{1}\left(1+|x(\xi)|+\left|P_{n} x(\xi)\right|\right)^{2 p_{2} p} d \xi\right)^{1 / 2} \\
& \times\left(\int_{0}^{1}\left|P_{n} x(\xi)-x(\xi)\right|^{2 p} d \xi\right)^{1 / 2}
\end{aligned}
$$

Since $x \mapsto\left\|1+|x|+\left|P_{n} x\right|\right\|_{L^{2 p_{2} p}(0,1)}$ is bounded in $L^{2 p_{2} p}(H, \mu)$ by a constant independent of $n$, and $\left\|P_{n} x-x\right\|_{L^{2 p}(0,1)}$ vanishes in $L^{2 p}(H, \mu)$ as $n \rightarrow \infty$, by the Hölder inequality the right-hand side vanishes in $L^{1}(H, \mu)$ as $n \rightarrow \infty$. Hence, $U$ in $L^{p}(H, \mu)$ and $U_{n} \rightarrow U$ in $L^{p}(H, \mu)$ as $n \rightarrow \infty$.

To prove that $U \in W_{1 / 2}^{1, p}(H, \mu)$ it is enough to show that the sequence $U_{n}$ is bounded in $W_{1 / 2}^{1, p}(H, \mu)$ (e.g., [3], Lemma 5.4.4). We already know that it is bounded in $L^{p}(H, \mu)$. Moreover each $U_{n}$ is continuously differentiable, since it is the composition of $x \mapsto P_{n} x$ which is smooth from $H$ to $C([0,1])$, and $y \mapsto \int_{0}^{1} \Phi(y(\xi)) d \xi$ which is continuously differentiable from $C([0,1])$ to $\mathbb{R}$, and

$$
\begin{equation*}
D_{k} U_{n}(x)=\int_{0}^{1} \Phi^{\prime}\left(P_{n} x(\xi)\right) f_{k}(\xi) d \xi, \quad k \leq n \tag{6.5}
\end{equation*}
$$

while $D_{k} U_{n}(x)=0$ for $k>n$. Using again assumption (5.4) and the Hölder inequality, we get

$$
\begin{aligned}
\left|D_{k} U_{n}(x)\right| & =\left|\int_{0}^{1} \Phi^{\prime}\left(P_{n} x(\xi)\right) f_{k}(\xi) d \xi\right| \leq C \int_{0}^{1}\left(1+\left|P_{n} x(\xi)\right|\right)^{p_{2}}\left|f_{k}(\xi)\right| d \xi \\
& \leq \frac{C}{\lambda_{k}^{1 / 4}}\left\|1+\left|P_{n} x\right|\right\|_{L^{2 p_{2}(0,1)}}^{p_{2}}
\end{aligned}
$$

for $k \leq n$. Then

$$
\left\|Q^{1 / 2} D U_{n}(x)\right\|^{2}=\sum_{k=1}^{n} \lambda_{k}\left|D_{k} U_{n}(x)\right|^{2} \leq C^{2} \sum_{k=1}^{\infty} \lambda_{k}^{1 / 2}\left\|1+\left|P_{n} x\right|\right\|_{L^{2 p_{2}(0,1)}}^{2 p_{2}}
$$

By the first part of the proof we know that $x \mapsto\left\|P_{n} x\right\|_{L^{2 p_{2}}(0,1)}^{2 p_{2}}$ belongs to $L^{1}(H, \mu)$ with norm bounded by a constant independent of $n$. Since $\sum_{k=1}^{\infty} \lambda_{k}^{1 / 2}<$ $\infty$, then $U_{n}$ is bounded in $W^{1, p}(H, \mu)$ so that $U \in W^{1, p}(H, \mu)$.

Now we show that for every $k \in \mathbb{N}$, a subsequence of $D_{k} U_{n}$ converges to $\int_{0}^{1} \Phi^{\prime}(x(\xi)) f_{k}(\xi) d \xi$ in $L^{2}(H, \mu)$. Then the equality $D_{k} U(x)=\int_{0}^{1} \Phi^{\prime}(x(\xi)) \times$ $f_{k}(\xi) d \xi \mu$-a.e. follows using the integration by parts formula (2.1).

We have

$$
\begin{aligned}
& \int_{H}\left|D_{k} U_{n}(x)-\int_{0}^{1} \Phi^{\prime}(x(\xi)) f_{k}(\xi) d \xi\right|^{2} d \mu \\
& \quad \leq \int_{H} \int_{0}^{1}\left|\Phi^{\prime}\left(P_{n} x(\xi)\right)-\Phi^{\prime}(x(\xi))\right|^{2} f_{k}(\xi)^{2} d \xi d \mu
\end{aligned}
$$

By Corollary 6.4, the sequence of functions $(x, \xi) \mapsto P_{n} x(\xi)$ converges to $x(\xi)$ in $L^{2}(H, \mu)$. Consequently, a subsequence converges $\mu$-almost everywhere, and since $\Phi^{\prime}$ is continuous, along such subsequence $(x, \xi) \mapsto\left(\Phi^{\prime}\left(P_{n} x(\xi)\right)-\right.$ $\left.\Phi^{\prime}(x(\xi))\right) f_{k}(\xi)$ vanishes. Moreover, by assumption (5.4),

$$
\left|\Phi^{\prime}\left(P_{n} x(\xi)\right)-\Phi^{\prime}(x(\xi))\right|^{2} f_{k}(\xi)^{2} \leq C^{2}\left(2+\left|P_{n} x(\xi)\right|^{p_{2}}+|x(\xi)|^{p_{2}}\right)\left\|f_{k}\right\|_{\infty}^{2}
$$

which belongs to $L^{1}(H \times(0,1), \mu \times d \xi)$ with norm bounded by a constant independent of $n$. The statement follows by the dominated convergence theorem.

Then, $U$ satisfies Hypothesis 2.1. So, the results of Theorem 3.9 and of Propositions 4.2, 4.6 hold.

We recall that the operator $Q^{1 / 2} D$ in the space $L^{2}(H, v ; H)$ is the closure of the operator $\varphi \mapsto Q^{1 / 2} D \varphi$ defined in a set of smooth functions; see Definition 2.5. However, we can identify $Q^{1 / 2} D U(x)$ : indeed, recalling that $B=Q^{-1 / 2} / \sqrt{2}$, we
obtain

$$
\begin{aligned}
D_{k} U(x) & =\left\langle\Phi^{\prime} \circ x, f_{k}\right\rangle_{L^{2}(0,1)}=\left\langle\Phi^{\prime} \circ x-\int_{0}^{1} \Phi^{\prime}(x(\xi)) d \xi, B f_{k}\right\rangle_{H} \\
& =\frac{\lambda_{k}^{-1 / 2}}{\sqrt{2}}\left\langle\Phi^{\prime} \circ x-\int_{0}^{1} \Phi^{\prime}(x(\xi)) d \xi, f_{k}\right\rangle_{H}
\end{aligned}
$$

for every $x \in \widetilde{L}^{2 p_{2}}(0,1)$, so that

$$
\begin{aligned}
Q^{1 / 2} D U(x) & =\frac{1}{\sqrt{2}} \sum_{k=1}^{\infty}\left\langle\Phi^{\prime} \circ x-\int_{0}^{1} \Phi^{\prime}(x(\xi)) d \xi, f_{k}\right\rangle_{H} f_{k} \\
& =\frac{\Phi^{\prime} \circ x-\int_{0}^{1} \Phi^{\prime}(x(\xi)) d \xi}{\sqrt{2}} .
\end{aligned}
$$

On the other hand, we already mentioned that if $\Phi^{\prime} \circ x \in V$ [i.e., $\left.\Phi^{\prime} \circ x \in D(B)\right]$, then $D_{0} U(x)=B\left(\Phi^{\prime} \circ x\right)$, so that, since $Q^{1 / 2}=B^{-1} / \sqrt{2}, Q^{1 / 2} D_{0} U(x)=$ $Q^{1 / 2} D U(x)$. For such $x$ we have

$$
\begin{aligned}
\left\langle B\left(\Phi^{\prime} \circ x\right), D u(x)\right\rangle & =\left\langle\Phi^{\prime} \circ x, B D u(x)\right\rangle=\left\langle Q^{1 / 2} D U(x), Q^{-1 / 2} D u(x)\right\rangle \\
& =\langle D U(x), D u(x)\rangle .
\end{aligned}
$$

Then the stochastic differential equation (1.2) in $H$ reads as

$$
\begin{equation*}
d X(t)=\left(-\frac{\partial^{4}}{\partial \xi^{4}} X-\frac{\partial^{2}}{\partial \xi^{2}} \Phi^{\prime}(X)\right) d t+d W(t), \quad X(0)=x \tag{6.6}
\end{equation*}
$$

and it is a stochastic Cahn-Hilliard equation, whose Kolmogorov operator is $\mathcal{K}$. It was studied in [16] and in several following papers, in particular in [9], where existence and uniqueness of weak solutions were proved for polynomial nonlinearities $\Phi$. Here $W(t)$ is, as usual, any $H$-valued cylindrical Wiener process defined in a probability space $(\Omega, \mathcal{F}, \mathbb{P})$.

We think that it is possible to relate the weak solution to (6.6) constructed in [9] to the solution of the Kolmogorov equation by formula (3.24), at least in the model case $\Phi(\xi)=\xi^{2 m}$ with $m \in \mathbb{N}$. Indeed, for every $x \in H$ the weak solution given by [9], Theorem 2.1, is obtained through cylindrical approximations $X_{n}(t)$, solutions to

$$
\begin{equation*}
d X_{n}=\left(A_{n} X_{n}+P_{n} B \Phi^{\prime}\left(P_{n} X\right)\right) d t+P_{n} d W(t), \quad X_{n}(0)=P_{n} x \tag{6.7}
\end{equation*}
$$

with $A_{n}=A_{\mid P_{n}(H)} \in \mathcal{L}\left(P_{n}(H)\right)$; identifying $P_{n}(H)$ with $\mathbb{R}^{n}$ the Kolmogorov operator $\mathcal{K}_{n}$ associated to (6.7) is

$$
\mathcal{K}_{n} \varphi=\frac{1}{2} \Delta \varphi-\sum_{k=1}^{n}\left(\frac{x_{k}}{2 \lambda_{k}}+\int_{0}^{1} \Phi^{\prime}\left(\sum_{h=1}^{n} x_{h} f_{h}(\xi)\right) f_{k}(\xi) d \xi\right) D_{k} \varphi
$$

Taking into account such explicit expressions, one should be able to follow the procedure of Proposition 3.8 (that deals with the case of Lipschitz continuous $D U$ ). However, many details should be fixed, and giving a complete proof goes beyond the aims of this paper.

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[^0]:    ${ }^{1}$ Note that $\rho$ may vanish on some set with positive measure, so that $f$ does not belong necessarily to $L^{\infty}(H, v)$, and even it does, its $L^{\infty}(H, v)$ norm may be bigger than its $L^{\infty}(H, \rho v)$ norm.

