# McKean-Vlasov SDEs with Drifts Discontinuous under Wasserstein Distance* 

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#### Abstract

Existence and uniqueness are proved for Mckean-Vlasov type distribution dependent SDEs with singular drifts satisfying an integrability condition in space variable and the Lipschitz condition in distribution variable with respect to $\mathbb{W}_{0}$ or $\mathbb{W}_{0}+\mathbb{W}_{\theta}$ for some $\theta \geq 1$, where $\mathbb{W}_{0}$ is the total variation distance and $\mathbb{W}_{\theta}$ is the $L^{\theta}$-Wasserstein distance. This improves some existing results where the drift is continuous in the distribution variable with respect to the Wasserstein distance.


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## 1 Introduction

Consider the following distribution dependent SDE on $\mathbb{R}^{d}$ :

$$
\begin{equation*}
\mathrm{d} X_{t}=b_{t}\left(X_{t}, \mathscr{L}_{X_{t}}\right) \mathrm{d} t+\sigma_{t}\left(X_{t}, \mathscr{L}_{X_{t}}\right) \mathrm{d} W_{t}, \quad t \in[0, T] \tag{1.1}
\end{equation*}
$$

where $T>0$ is a fixed time, $\left(W_{t}\right)_{t \in[0, T]}$ is the $m$-dimensional Brownian motion on a complete filtration probability space $\left(\Omega,\left\{\mathscr{F}_{t}\right\}_{t \in[0, T]}, \mathbb{P}\right), \mathscr{L}_{X_{t}}$ is the law of $X_{t}$,

$$
b:[0, T] \times \mathbb{R}^{d} \times \mathscr{P} \rightarrow \mathbb{R}^{d}, \quad \sigma:[0, T] \times \mathbb{R}^{d} \times \mathscr{P} \rightarrow \mathbb{R}^{d} \otimes \mathbb{R}^{m}
$$

[^0]are measurable, and $\mathscr{P}$ is the space of all probability measures on $\mathbb{R}^{d}$ equipped with the weak topology.

This type SDEs are also called McKean-Vlasov SDEs and mean field SDEs, and have been intensively investigated due to its wide applications, see for instance $[1,2,5,8,10$, $11,12,20,22]$ and references within.

An adapted continuous process on $\mathbb{R}^{d}$ is called a (strong) solution of (1.1), if

$$
\begin{equation*}
\mathbb{E} \int_{0}^{T}\left\{\left|b_{t}\left(X_{t}, \mathscr{L}_{X_{t}}\right)\right|+\left\|\sigma_{t}\left(X_{t}, \mathscr{L}_{X_{t}}\right)\right\|^{2}\right\} \mathrm{d} t<\infty \tag{1.2}
\end{equation*}
$$

and $\mathbb{P}$-a.s.

$$
\begin{equation*}
X_{t}=X_{0}+\int_{0}^{t} b_{s}\left(X_{s}, \mathscr{L}_{X_{s}}\right) \mathrm{d} s+\int_{0}^{t} \sigma_{s}\left(X_{s}, \mathscr{L}_{X_{s}}\right) \mathrm{d} W_{s}, \quad t \in[0, T] . \tag{1.3}
\end{equation*}
$$

We call (1.1) (strongly) well-posed for an $\mathscr{F}_{0}$-measurable initial value $X_{0}$, if (1.1) has a unique solution starting at $X_{0}$.

When a different probability measure $\tilde{\mathbb{P}}$ is concerned, we use $\mathscr{L}_{\xi} \mid \tilde{\mathbb{P}}$ to denote the law of a random variable $\xi$ under the probability $\tilde{\mathbb{P}}$, and use $\mathbb{E}_{\tilde{\mathbb{P}}}$ to stand for the expectation under $\tilde{\mathbb{P}}$. For any $\mu_{0} \in \mathscr{P},\left(\tilde{X}_{t}, \tilde{W}_{t}\right)_{t \in[0, T]}$ is called a weak solution to (1.1) starting at $\mu_{0}$, if $\left(\tilde{W}_{t}\right)_{t \in[0, T]}$ is the $m$-dimensional Brownian motion under a complete filtration probability space $\left(\tilde{\Omega},\left\{\tilde{\mathscr{F}}_{t}\right\}_{t \in[0, T]}, \tilde{\mathbb{P}}\right),\left(\tilde{X}_{t}\right)_{t \in[0, T]}$ is a continuous $\tilde{\mathscr{F}}_{t}$-adapted process on $\mathbb{R}^{d}$ with $\mathscr{L}_{\tilde{X}_{0}} \mid \tilde{\mathbb{P}}=\mu_{0}$, and (1.2)-(1.3) hold for $\left(\tilde{X}, \tilde{W}, \tilde{\mathbb{P}}, \mathbb{E}_{\tilde{P}}\right)$ replacing $(X, W, \mathbb{P}, \mathbb{E})$. We call (1.1) weakly well-posed for an initial distribution $\mu_{0}$, if it has a unique weak solution starting at $\mu_{0}$; i.e. it has a weak solution $\left(\tilde{X}_{t}, \tilde{W}_{t}\right)_{t \in[0, T]}$ with initial distribution $\mu_{0}$ under some complete filtration probability space $\left(\tilde{\Omega},\left\{\tilde{\mathscr{F}}_{t}\right\}_{t \in[0, T]}, \tilde{\mathbb{P}}\right)$, and $\mathscr{L}_{\tilde{X}_{[0, T]}}\left|\tilde{\mathbb{P}}^{2}=\mathscr{L}_{\bar{X}_{[0, T]}}\right| \overline{\mathbb{P}}$ holds for any other weak solution with the same initial distribution $\left(\bar{X}_{t}, \bar{W}_{t}\right)_{t \in[0, T]}$ under some complete filtration probability space $\left(\bar{\Omega},\left\{\overline{\mathscr{F}}_{t}\right\}_{t \in[0, T]}, \overline{\mathbb{P}}\right)$.

Recently, the (weak and strong) well-posedness is studied in [3, 4, 6, 13, 16, 17, 19] for (1.1) with $\sigma_{t}(x, \gamma)=\sigma_{t}(x)$ independent of the distribution variable $\gamma$, and with singular drift $b_{t}(x, \gamma)$. See also $[12,16]$ for the case with memory. We briefly recall some conditions on $b$ which together with a regular and non-degenerate condition on $\sigma$ implies the wellposedness of (1.1). To this end, we recall the $L^{\theta}$-Wasserstein distance $\mathbb{W}_{\theta}$ for $\theta>0$ :

$$
\mathbb{W}_{\theta}(\gamma, \tilde{\gamma}):=\inf _{\pi \in \mathscr{C}(\gamma, \tilde{\gamma})}\left(\int_{\mathbb{R}^{d} \times \mathbb{R}^{d}}|x-y|^{\theta} \pi(\mathrm{d} x, \mathrm{~d} y)\right)^{\frac{1}{1 \mathrm{1v} \theta}}, \quad \gamma, \tilde{\gamma} \in \mathscr{P}
$$

where $\mathscr{C}(\gamma, \tilde{\gamma})$ is the set of all couplings of $\gamma$ and $\tilde{\gamma}$. By the convention that $r^{0}=1_{\{r>0\}}$ for $r \geq 0$, we may regard $\mathbb{W}_{0}$ as the total variation distance, i.e. set

$$
\mathbb{W}_{0}(\gamma, \tilde{\gamma})=\|\gamma-\tilde{\gamma}\|_{T V}:=\sup _{A \in \mathscr{B}\left(\mathbb{R}^{d}\right)}|\gamma(A)-\tilde{\gamma}(A)| .
$$

References [3, 4] give the well-posedness of (1.1) with a deterministic initial value $X_{0} \in \mathbb{R}^{d}$, where the drift $b_{t}(x, \gamma)$ is assumed to be linear growth in $x$ uniformly in $t, \gamma$,
and

$$
\left|b_{t}(x, \gamma)-b_{t}(x, \tilde{\gamma})\right| \leq \phi\left(\mathbb{W}_{1}(\gamma, \tilde{\gamma})\right)
$$

holds for some function $\phi \in C((0, \infty) ;(0, \infty))$ with $\int_{0} \frac{1}{\phi(s)} \mathrm{d} s=\infty$. Note that for distribution dependent SDEs the well-posedness for deterministic initial values does not imply that for random ones.
[17, Theorem 3] presents the well-posedness of (1.1) with exponentially integrable $X_{0}$ and a drift $b$ of type

$$
\begin{equation*}
b_{t}(x, \gamma):=\int_{\mathbb{R}^{d}} \tilde{b}_{t}(x, y) \gamma(\mathrm{d} y), \tag{1.4}
\end{equation*}
$$

where $\tilde{b}_{t}(x, y)$ has linear growth in $x$ uniformly in $t$ and $y$. Since $\tilde{b}_{t}(x, y)$ is bounded in $y$, $b_{t}(x, \cdot)$ is Lipschtiz continuous in the total variation distance $\mathbb{W}_{0}$. [19] considers the same type drift and proves the well-posedness of (1.1) under the conditions that $\mathbb{E}\left|X_{0}\right|^{\beta}<\infty$ for some $\beta>0$ and

$$
\left|\tilde{b}_{t}(x, y)\right| \leq h_{t}(x-y)
$$

for some $h \in L^{q}\left([0, T] ; \tilde{L}^{p}\left(\mathbb{R}^{d}\right)\right)$ for some $p, q>1$ with $\frac{d}{p}+\frac{2}{q}<1$, where $\tilde{L}^{p}$ is a localized $L^{p}$ space.

In [6] the well-posedness of (1.1) is proved for $X_{0}$ satisfying $\mathbb{E}\left|X_{0}\right|^{2}<\infty$, and for $b$ given by

$$
\begin{equation*}
b_{t}(x, \gamma)=\tilde{b}_{t}(x, \gamma(\varphi)), \tag{1.5}
\end{equation*}
$$

where $\gamma(\varphi):=\int_{\mathbb{R}^{d}} \varphi \mathrm{~d} \gamma$ for some $\alpha$-Hölder continuous function $\varphi$, and $\left|\tilde{b}_{t}(x, r)\right|+\left|\partial_{r} \tilde{b}_{t}(x, r)\right|$ is bounded. Consequently, $b_{t}(x, \gamma)$ is bounded and Lipschitz continuous in $\gamma$ with respect to $\mathbb{W}_{\alpha}$.

In [13] the well-posedness is derived under the conditions that $\mathbb{E}\left|X_{0}\right|^{\theta}<\infty$ for some $\theta \geq$ $1, b_{t}(x, \gamma)$ is Lipschitz continuous in $\gamma$ with respect to $\mathbb{W}_{\theta}$, and for any $\mu \in C\left([0, T] ; \mathscr{P}_{\theta}\right)$,

$$
b_{t}^{\mu}(x):=b_{t}\left(x, \mu_{t}\right), \quad(t, x) \in[0, T] \times \mathbb{R}^{d}
$$

satisfies $\left|b^{\mu}\right|^{2} \in L_{p}^{q}(T)$ for some $(p, q) \in \mathscr{K}$, where

$$
\begin{aligned}
& L_{p}^{q}(T):=\left\{f \in \mathscr{B}\left([0, T] \times \mathbb{R}^{d}\right): \int_{0}^{T}\left(\int_{\mathbb{R}^{d}}\left|f_{t}(x)\right|^{p} \mathrm{~d} x\right)^{\frac{q}{p}} \mathrm{~d} t<\infty\right\}, \\
& \mathscr{K}:=\left\{(p, q) \in(1, \infty) \times(1, \infty): \frac{d}{p}+\frac{2}{q}<2\right\} .
\end{aligned}
$$

Moreover, in [15] the well-posedness of (1.1) has been proved for

$$
\begin{equation*}
b_{t}(x, \mu)=\tilde{b}\left(\rho_{\mu}(x)\right), \quad \sigma_{t}(x, \mu)=\tilde{\sigma}\left(\rho_{\mu}(x)\right) \tag{1.6}
\end{equation*}
$$

with initial distribution having density function (with respect to the Lebesgue mseaure) in the class $H^{2+\alpha}$ for some $\alpha>0$, where $\rho_{\mu}$ is the density function of $\mu$ with respect to the

Lebesgue measure, $\tilde{b} \in C^{2}\left([0, \infty) ; \mathbb{R}^{d}\right)$ and $\tilde{\sigma} \in C^{3}\left([0, \infty) ; \mathbb{R}^{d} \otimes \mathbb{R}^{d}\right)$. As for the weak wellposedness, [14] assumes that $b$ is bounded and $\mathbb{W}_{0}$-Lipschitz continuous in distribution variable, and $\sigma$ is Lipschitz continuous in space variable.

In this paper, we prove the (weak and strong) well-posedness of (1.1) for general type $b$ with $b_{t}(x, \gamma)$ Lipschitz continuous in $\gamma$ under the metric $\mathbb{W}_{0}$ or $\mathbb{W}_{0}+\mathbb{W}_{\theta}$ for some $\theta \geq 1$. This condition is weaker than those in [3, 4, 6, 13] in the sense that the drift is not necessarily continuous in the Wasserstein distance, but is incomparable with those in $[17,19]$ where $b$ is of the integral type as in (1.4). Moreover, our result works for any initial value and initial distribution.

Recall that a continuous function $f$ on $\mathbb{R}^{d}$ is called weakly differentiable, if there exists (hence unique) $\xi \in L_{l o c}^{1}\left(\mathbb{R}^{d}\right)$ such that

$$
\int_{\mathbb{R}^{d}}(f \Delta g)(x) \mathrm{d} x=-\int_{\mathbb{R}^{d}}\langle\xi, \nabla g\rangle(x) \mathrm{d} x, \quad g \in C_{0}^{\infty}\left(\mathbb{R}^{d}\right) .
$$

In this case, we write $\xi=\nabla f$ and call it the weak gradient of $f$. For $p, q \geq 1$, let
$L_{p, l o c}^{q}(T)=\left\{f \in \mathscr{B}\left([0, T] \times \mathbb{R}^{d}\right): \int_{0}^{T}\left(\int_{K}\left|f_{t}(x)\right|^{p} \mathrm{~d} x\right)^{\frac{q}{p}} \mathrm{~d} t<\infty, K \subset \mathbb{R}^{d}\right.$ is compact $\}$.
We will use the following conditions.
$\left(A_{\sigma}\right) \sigma_{t}(x, \gamma)=\sigma_{t}(x)$ is uniformly continuous in $x \in \mathbb{R}^{d}$ uniformly in $t \in[0, T]$; the weak gradient $\nabla \sigma_{t}$ exists for a.e. $t \in[0, T]$ such that $|\nabla \sigma|^{2} \in L_{p}^{q}(T)$ for some $(p, q) \in \mathscr{K}$; and there exists a constant $K_{1} \geq 1$ such that

$$
\begin{equation*}
K_{1}^{-1} I \leq\left(\sigma_{t} \sigma_{t}^{*}\right)(x) \leq K_{1} I, \quad(t, x) \in[0, T] \times \mathbb{R}^{d} \tag{1.7}
\end{equation*}
$$

where $I$ is the $d \times d$ identity matrix.
$\left(A_{b}\right) b=\bar{b}+\hat{b}$, where $\bar{b}$ and $\hat{b}$ satisfy

$$
\begin{align*}
& \left|\hat{b}_{t}(x, \gamma)-\hat{b}_{t}(y, \tilde{\gamma})\right|+\left|\bar{b}_{t}(x, \gamma)-\bar{b}_{t}(x, \tilde{\gamma})\right| \\
& \quad \leq K_{2}\left(\|\gamma-\tilde{\gamma}\|_{T V}+\mathbb{W}_{\theta}(\gamma, \tilde{\gamma})+|x-y|\right), \quad t \in[0, T], x, y \in \mathbb{R}^{d}, \gamma, \tilde{\gamma} \in \mathscr{P}_{\theta} \tag{1.8}
\end{align*}
$$

for some constants $\theta, K_{2} \geq 1$, and there exists $(p, q) \in \mathscr{K}$ such that

$$
\begin{equation*}
\sup _{t \in[0, T]}\left|\hat{b}_{t}\left(0, \delta_{0}\right)\right|+\sup _{\mu \in C\left([0, T] ; \mathscr{P}_{\theta}\right)}\left\|\left|\bar{b}^{\mu}\right|^{2}\right\|_{L_{p}^{q}(T)}<\infty \tag{1.9}
\end{equation*}
$$

where $\bar{b}_{t}^{\mu}(x):=\bar{b}_{t}\left(x, \mu_{t}\right)$ for $(t, x) \in[0, T] \times \mathbb{R}^{d}$, and $\delta_{0}$ stands for the Dirac measure at the point $0 \in \mathbb{R}^{d}$.
$\left(A_{b}^{\prime}\right)$ For any $\mu \in \mathscr{B}([0, T] ; \mathscr{P}),\left|b^{\mu}\right|^{2} \in L_{p, l o c}^{q}(T)$ for some $(p, q) \in \mathscr{K}$. Moreover, there exists a function $\Gamma:[0, \infty) \rightarrow[0, \infty)$ satisfying $\int_{1}^{\infty} \frac{1}{\Gamma(x)}=\infty$ such that

$$
\begin{equation*}
\left\langle b_{t}\left(x, \delta_{0}\right), x\right\rangle \leq \Gamma\left(|x|^{2}\right), \quad t \in[0, T], x \in \mathbb{R}^{d} . \tag{1.10}
\end{equation*}
$$

In addition, there exists a constant $K_{3} \geq 1$ such that

$$
\begin{equation*}
\left|b_{t}(x, \gamma)-b_{t}(x, \tilde{\gamma})\right| \leq K_{3}\|\gamma-\tilde{\gamma}\|_{T V}, \quad t \in[0, T], x \in \mathbb{R}^{d}, \gamma, \tilde{\gamma} \in \mathscr{P} . \tag{1.11}
\end{equation*}
$$

When (1.1) is weakly well-posed for initial distribution $\gamma$, we denote $P_{t}^{*} \gamma$ the distribution of the weak solution at time $t$.

Theorem 1.1. Assume $\left(A_{\sigma}\right)$.
(1) If $\left(A_{b}^{\prime}\right)$ holds, then (1.1) is strongly and weakly well-posed for any initial values and any initial distribution. Moreover,

$$
\begin{equation*}
\left\|P_{t}^{*} \mu_{0}-P_{t}^{*} \nu_{0}\right\|_{T V}^{2} \leq 2 \mathrm{e}^{\frac{K_{1} K_{3}^{2} t}{2}}\left\|\mu_{0}-\nu_{0}\right\|_{T V}^{2}, \quad t \in[0, T], \mu_{0}, \nu_{0} \in \mathscr{P} . \tag{1.12}
\end{equation*}
$$

(2) Let $\mathbb{E}\left|X_{0}\right|^{\theta}<\infty$ and $\mu_{0}\left(|\cdot|^{\theta}\right)<\infty$. If $\left(A_{b}\right)$ holds, then (1.1) is strongly well-posed for initial value $X_{0}$ and weakly well-posed for initial distribution $\mu_{0}$. Moreover, there exists a constant $c>0$ such that for any $\mu_{0}, \nu_{0} \in \mathscr{P}_{\theta}$,

$$
\begin{align*}
& \left\|P_{t}^{*} \mu_{0}-P_{t}^{*} \nu_{0}\right\|_{T V}+\mathbb{W}_{\theta}\left(P_{t}^{*} \mu_{0}, P_{t}^{*} \nu_{0}\right)  \tag{1.13}\\
& \leq c\left\{\left\|\mu_{0}-\nu_{0}\right\|_{T V}+W_{\theta}\left(\mu_{0}, \nu_{0}\right)\right\}, \quad t \in[0, T] .
\end{align*}
$$

To illustrate this result comparing with earlier ones, we present an example of $b$ which satisfies our conditions but is not of type (1.4)-(1.6) and is discontinuous in both the space variable and the distribution variable under the weak topology. If one wants to control a stochastic system in terms of an ideal reference distribution $\mu_{0}$, it is natural to take a drift depending on a probability distance between $\mu_{0}$ and the law of the system. As two typical probability distances, the total variation and Wasserstein distances have been widely applied in applications. So, we take for instance

$$
b_{t}(x, \mu)=\bar{b}(t, x, \mu)+h\left(t, x, \mathbb{W}_{\theta}\left(\mu, \mu_{0}\right),\left\|\mu-\mu_{0}\right\|_{T V}\right)
$$

for some $\theta \geq 1$, where $\bar{b}$ satisfies (1.8) and (1.9) for $\hat{b}=0$ which refers to the singularity in the space variable $x$, and $h:[0, T] \times \mathbb{R}^{d} \times[0, \infty)^{2} \rightarrow \mathbb{R}^{d}$ is measurable such that $h(t, x, r, s)$ is bounded in $t \in[0, T]$ and Lipschitz continuous in $(x, r, s) \in \mathbb{R}^{d} \times[0, \infty)^{2}$ uniformly in $t \in[0, T]$. Obviously, $b(t, x, \mu)$ satisfies condition $\left(A_{b}\right)$ but is not of type (1.4)-(1.6) and can be discontinuous in $x$ and $\mu$ under the weak topology.

In the next section we make some preparations, which will be used in Section 3 for the proof of Theorem 1.1.

## 2 Preparations

We first present the following version of Yamada-Watanabe principle modified from [13, Lemma 3.4].
Lemma 2.1. Assume that (1.1) has a weak solution $\left(\bar{X}_{t}\right)_{t \in[0, T]}$ under probability $\overline{\mathbb{P}}$, and let $\mu_{t}=\mathscr{L}_{\bar{X}_{t}} \mid \overline{\mathbb{P}}, t \in[0, T]$. If the $S D E$

$$
\begin{equation*}
\mathrm{d} X_{t}=b_{t}\left(X_{t}, \mu_{t}\right) \mathrm{d} t+\sigma_{t}\left(X_{t}, \mu_{t}\right) \mathrm{d} W_{t} \tag{2.1}
\end{equation*}
$$

has strong uniqueness for some initial value $X_{0}$ with $\mathscr{L}_{X_{0}}=\mu_{0}$, then (1.1) has a strong solution starting at $X_{0}$. If moreover (1.1) has strong uniqueness for any initial value $X_{0}$ with $\mathscr{L}_{X_{0}}=\mu_{0}$, then it is weakly well-posed for the initial distribution $\mu_{0}$.
Proof. (a) Strong existence. Since $\mu_{t}=\mathscr{L}_{\bar{X}_{t}} \mid \overline{\mathbb{P}}, \bar{X}_{t}$ under $\overline{\mathbb{P}}$ is also a weak solution of (2.1) with initial distribution $\mu_{0}$. By the Yamada-Watanabe principle, the strong uniqueness of (2.1) with initial value $X_{0}$ implies the strong (resp. weak) well-posedness of (2.1) starting at $X_{0}$ (resp. $\mu_{0}$ ). In particular, the weak uniqueness implies $\mathscr{L}_{X_{t}}=\mu_{t}, t \in[0, T]$, so that $X_{t}$ solves (1.1).
(b) Weak uniqueness. Let $\tilde{X}_{t}$ under probability $\tilde{\mathbb{P}}$ be another weak solution of (1.1) with initial distribution $\mu_{0}$. For any initial value $X_{0}$ with $\mathscr{L}_{X_{0}}=\mu_{0}$, the strong uniqueness of (2.1) starting at $X_{0}$ implies

$$
X_{[0, T]}=F\left(X_{0}, W_{[0, T]}\right)
$$

for some measurable function $F: \mathbb{R}^{d} \times C\left([0, T] ; \mathbb{R}^{d}\right) \rightarrow C\left([0, T] ; \mathbb{R}^{d}\right)$. This and the weak uniqueness of (2.1) proved in (a) yield

$$
\begin{equation*}
\mathscr{L}_{\bar{X}_{[0, T]}}\left|\overline{\mathbb{P}}=\mathscr{L}_{X_{[0, T]}}\right| \mathbb{P} \tag{2.2}
\end{equation*}
$$

Let $\hat{X}_{[0, T]}=F\left(\tilde{X}_{0}, \tilde{W}_{[0, T]}\right)$. We have $\hat{X}_{0}=\tilde{X}_{0}$ and

$$
\mathscr{L}_{\hat{X}_{[0, T]} \mid}\left|\tilde{\mathbb{P}}=\mathscr{L}_{X_{[0, T]}}\right| \mathbb{P}
$$

This and (2.2) imply $\mathscr{L}_{\hat{X}_{t}} \mid \tilde{\mathbb{P}}=\mu_{t}$, so that $\hat{X}_{t}$ under $\tilde{\mathbb{P}}$ is a weak solution of (1.1) with $\hat{X}_{0}=\tilde{X}_{0}$. By the strong uniqueness of (1.1), we derive $\hat{X}_{[0, T]}=\tilde{X}_{[0, T]}$. Combining this with (2.2) we obtain

$$
\mathscr{L}_{\tilde{X}_{[0, T]}}\left|\tilde{\mathbb{P}}=\mathscr{L}_{\hat{X}_{[0, T]}}\right| \tilde{\mathbb{P}}=\mathscr{L}_{X_{[0, T]}}\left|\mathbb{P}=\mathscr{L}_{\bar{X}_{[0, T]}}\right| \overline{\mathbb{P}}
$$

i.e. (1.1) has weak uniqueness starting at $\mu_{0}$.

We will use the following result for the maximal operator:

$$
\begin{equation*}
\mathscr{M} h(x):=\sup _{r>0} \frac{1}{|B(x, r)|} \int_{B(x, r)} h(y) \mathrm{d} y, \quad h \in L_{l o c}^{1}\left(\mathbb{R}^{d}\right), x \in \mathbb{R}^{d}, \tag{2.3}
\end{equation*}
$$

where $B(x, r):=\{y:|x-y|<r\}$, see [7, Appendix A].

Lemma 2.2. There exists a constant $C>0$ such that for any continuous and weak differentiable function $f$,

$$
\begin{equation*}
|f(x)-f(y)| \leq C|x-y|(\mathscr{M}|\nabla f|(x)+\mathscr{M}|\nabla f|(y)), \text { a.e. } x, y \in \mathbb{R}^{d} \tag{2.4}
\end{equation*}
$$

Moreover, for any $p>1$, there exists a constant $C_{p}>0$ such that

$$
\begin{equation*}
\|\mathscr{M} f\|_{L^{p}} \leq C_{p}\|f\|_{L^{p}}, \quad f \in L^{p}\left(\mathbb{R}^{d}\right) . \tag{2.5}
\end{equation*}
$$

To compare the distribution dependent $\operatorname{SDE}$ (1.1) with a classical one, for any $\mu \in$ $\mathscr{B}([0, T] ; \mathscr{P})$, let $b_{t}^{\mu}(x):=b_{t}\left(x, \mu_{t}\right)$ and consider the classical SDE

$$
\begin{equation*}
\mathrm{d} X_{t}^{\mu}=b_{t}^{\mu}\left(X_{t}^{\mu}\right) \mathrm{d} t+\sigma_{t}\left(X_{t}^{\mu}\right) \mathrm{d} W_{t}, \quad t \in[0, T] . \tag{2.6}
\end{equation*}
$$

According to [25], assumption $\left(A_{\sigma}\right)$ together with $\left(A_{b}\right)$ or $\left(A_{b}^{\prime}\right)$ implies the strong wellposedness, where under $\left(A_{b}^{\prime}\right)$ the non-explosion is implied by (1.10). For any $\gamma \in \mathscr{P}$, Let $\Phi_{t}^{\gamma}(\mu)=\mathscr{L}_{X_{t}^{\mu}}$ for $\left(X_{t}^{\mu}\right)_{t \in[0, T]}$ solving (2.6) with $\mathscr{L}_{X_{0}^{\mu}}=\gamma$. We have the following result.

Lemma 2.3. Assume $\left(A_{\sigma}\right)$ and let $\gamma \in \mathscr{P}$.
(1) If ( $A_{b}^{\prime}$ ) holds, then for any $\mu, \nu \in \mathscr{B}([0, T]$; $\mathscr{P})$,

$$
\begin{equation*}
\left\|\Phi_{t}^{\gamma}(\mu)-\Phi_{t}^{\gamma}(\nu)\right\|_{T V}^{2} \leq \frac{K_{1} K_{3}^{2}}{4} \int_{0}^{t}\left\|\mu_{s}-\nu_{s}\right\|_{T V}^{2} \mathrm{~d} s, \quad t \in[0, T] . \tag{2.7}
\end{equation*}
$$

(2) If $\left(A_{b}\right)$ holds and $\gamma \in \mathscr{P}_{\theta}$, then for any $\mu \in C\left([0, T] ; \mathscr{P}_{\theta}\right)$, we have $\Phi^{\gamma}(\mu) \in$ $C\left([0, T] ; \mathscr{P}_{\theta}\right)$. Moreover, for any $m \geq 1 \vee \frac{\theta}{2}$, there exists a constant $C>0$ such that for any $\mu, \nu \in C\left([0, T] ; \mathscr{P}_{\theta}\right)$ and $\gamma_{1}, \gamma_{2} \in \mathscr{P}_{\theta}$,

$$
\begin{align*}
& \left\{\mathbb{W}_{\theta}\left(\Phi_{t}^{\gamma_{1}}(\mu), \Phi_{t}^{\gamma_{2}}(\nu)\right)\right\}^{2 m} \\
& \leq C \mathbb{W}_{\theta}\left(\gamma_{1}, \gamma_{2}\right)^{2 m}+C \int_{0}^{t}\left\{\left\|\mu_{s}-\nu_{s}\right\|_{T V}+\mathbb{W}_{\theta}\left(\mu_{s}, \nu_{s}\right)\right\}^{2 m} \mathrm{~d} s, \quad t \in[0, T] \tag{2.8}
\end{align*}
$$

Proof. (1) Let $\left(A_{b}^{\prime}\right)$ hold and take $\mu, \nu \in \mathscr{B}([0, T] ; \mathscr{P})$. To compare $\Phi_{t}^{\gamma}(\mu)$ with $\Phi_{t}^{\gamma}(\nu)$, we rewrite (2.6) as

$$
\begin{equation*}
\mathrm{d} X_{t}^{\mu}=b_{t}\left(X_{t}^{\mu}, \nu_{t}\right) \mathrm{d} t+\sigma_{t}\left(X_{t}^{\mu}\right) \mathrm{d} \tilde{W}_{t} \tag{2.9}
\end{equation*}
$$

where

$$
\tilde{W}_{t}=W_{t}+\int_{0}^{t} \xi_{s} \mathrm{~d} s, \quad \xi_{s}:=\left\{\sigma_{s}^{*}\left(\sigma_{s} \sigma_{s}^{*}\right)^{-1}\right\}\left(X_{s}^{\mu}\right)\left[b_{s}\left(X_{s}^{\mu}, \mu_{s}\right)-b_{s}\left(X_{s}^{\mu}, \nu_{s}\right)\right], \quad s, t \in[0, T] .
$$

Noting that (1.7) together with (1.11) implies

$$
\begin{equation*}
\mathbb{E}\left[\mathrm{e}^{\frac{1}{2} \int_{0}^{T}\left|\xi_{s}\right|^{2} \mathrm{~d} s}\right]<\infty \tag{2.10}
\end{equation*}
$$

by the Girsanov theorem we see that $R_{T}:=\mathrm{e}^{-\int_{0}^{T}\left\langle\xi_{s}, \mathrm{~d} W_{s}\right\rangle-\frac{1}{2} \int_{0}^{T}\left|\xi_{s}\right|^{2} \mathrm{~d} s}$ is a probability density with respect to $\mathbb{P}$, and $\left(\tilde{W}_{t}\right)_{t \in[0, T]}$ is a $d$-dimensional Brownian motion under the probability $\mathbb{Q}:=R_{T} \mathbb{P}$.

By the weak uniqueness of (2.6) and $\mathscr{L}_{X_{0}^{\mu}} \mid \mathbb{Q}=\mathscr{L}_{X_{0}^{\mu}}=\gamma$, we conclude from (2.9) with $\mathbb{Q}$-Brownian motion $\tilde{W}_{t}$ that

$$
\Phi_{t}^{\gamma}(\nu)=\mathscr{L}_{X_{t}^{\mu}} \mid \mathbb{Q}, \quad t \in[0, T] .
$$

Combining this with $\left(A_{\sigma}\right)$ and applying Pinker's inequality [18], we obtain

$$
\begin{align*}
& 2\left\|\Phi_{t}^{\gamma}(\nu)-\Phi_{t}^{\gamma}(\mu)\right\|_{T V}^{2} \leq 2 \sup _{\|f\|_{\infty} \leq 1}\left(\mathbb{E}\left|f\left(X_{t}^{\mu}\right)\left(R_{t}-1\right)\right|\right)^{2}=2\left(\mathbb{E}\left|R_{t}-1\right|\right)^{2} \\
& \leq \mathbb{E}\left[R_{t} \log R_{t}\right]=\frac{1}{2} \mathbb{E}_{\mathbb{Q}} \int_{0}^{t}\left|\xi_{s}\right|^{2} \mathrm{~d} s  \tag{2.11}\\
& \leq \frac{K_{1}}{2} \mathbb{E}_{\mathbb{Q}} \int_{0}^{t}\left|b_{s}\left(X_{s}^{\mu}, \mu_{s}\right)-b_{s}\left(X_{s}^{\mu}, \nu_{s}\right)\right|^{2} \mathrm{~d} s
\end{align*}
$$

By $\left(A_{b}^{\prime}\right)$, this implies (2.7).
(2) Let $\left(A_{b}\right)$ hold and take $m \geq 1 \vee \frac{\theta}{2}$. Take $\mathscr{F}_{0}$-measurable random variables $X_{0}^{\mu}$ and $X_{0}^{\nu}$ such that $\mathscr{L}_{X_{0}^{\mu}}=\gamma_{1}, \mathscr{L}_{X_{0}^{\nu}}=\gamma_{2}$ and

$$
\mathbb{E}\left|X_{0}^{\mu}-X_{0}^{\nu}\right|^{\theta}=\left\{\mathbb{W}_{\theta}\left(\gamma_{1}, \gamma_{2}\right)\right\}^{\theta}
$$

Let $X_{t}^{\mu}$ solve (2.6) and $X_{t}^{\nu}$ solve the same SDE for $\nu$ replacing $\mu$. We need to find a constant $C>0$ such that for any $t \in[0, T]$,

$$
\begin{align*}
& \left\{\mathbb{W}_{\theta}\left(\Phi_{t}^{\gamma_{1}}(\mu), \Phi_{t}^{\gamma_{2}}(\nu)\right)\right\}^{2 m} \\
& \leq C\left(\mathbb{E}\left|X_{0}^{\mu}-X_{0}^{\nu}\right|^{\theta}\right)^{\frac{2 m}{\theta}}+C \int_{0}^{t}\left(\mathbb{W}_{\theta}\left(\mu_{s}, \nu_{s}\right)+\left\|\mu_{s}-\nu_{s}\right\|_{T V}\right)^{2 m} \mathrm{~d} s, \quad t \in[0, T] \tag{2.12}
\end{align*}
$$

To this end, we make a Zvokin type transform as in [13] and [24].
For any $\lambda>0$, consider the following PDE for $u:[0, T] \times \mathbb{R}^{d} \rightarrow \mathbb{R}^{d}$ :

$$
\begin{equation*}
\frac{\partial u_{t}}{\partial t}+\frac{1}{2} \operatorname{Tr}\left(\sigma_{t} \sigma_{t}^{*} \nabla^{2} u_{t}\right)+\nabla_{b_{t}^{\mu}} u_{t}+\bar{b}_{t}^{\mu}=\lambda u_{t}, \quad u_{T}=0 \tag{2.13}
\end{equation*}
$$

According to [24, Remark 2.1, Proposition 2.3 (2)], under assumptions $\left(A_{\sigma}\right)$ and $\left(A_{b}\right)$, when $\lambda$ is large enough (2.13) has a unique solution $\mathbf{u}^{\lambda, \mu}$ satisfying

$$
\begin{equation*}
\left\|\mathbf{u}^{\lambda, \mu}\right\|_{\infty}+\left\|\nabla \mathbf{u}^{\lambda, \mu}\right\|_{\infty} \leq \frac{1}{5} \tag{2.14}
\end{equation*}
$$

and

$$
\begin{equation*}
\left\|\nabla^{2} \mathbf{u}^{\lambda, \mu}\right\|_{L_{2 p}^{2 q}(T)}<\infty \tag{2.15}
\end{equation*}
$$

Let $\Theta_{t}^{\lambda, \mu}(x)=x+\mathbf{u}_{t}^{\lambda, \mu}(x)$. It is easy to see that (2.13) and the Itô formula imply

$$
\begin{equation*}
\mathrm{d} \Theta_{t}^{\lambda, \mu}\left(X_{t}^{\mu}\right)=\left(\lambda \mathbf{u}_{t}^{\lambda, \mu}+\hat{b}_{t}^{\mu}\right)\left(X_{t}^{\mu}\right) \mathrm{d} t+\left(\left\{\nabla \Theta_{t}^{\lambda, \mu}\right\} \sigma_{t}\right)\left(X_{t}^{\mu}\right) \mathrm{d} W_{t} . \tag{2.16}
\end{equation*}
$$

In particular, (2.14) and $\mathbb{E}\left[\left|X_{0}^{\mu}\right|^{\theta}\right]<\infty$ imply that $\mathbb{E}\left[\left|\Theta_{0}^{\lambda, \mu}\left(X_{0}^{\mu}\right)\right|^{\theta}\right]<\infty$ and (2.16) is an SDE for $\xi_{t}:=\Theta_{t}^{\lambda, \mu}\left(X_{t}^{\mu}\right)$ with coefficients of at most linear growth, so that $\mathscr{L}_{\xi}$. $\in$ $C\left([0, T] ; \mathscr{P}_{\theta}\right)$ and so does $\mathscr{L}_{X^{\mu}}$ due to (2.14).

It remains to prove (2.8). To this end, we observe that (2.13) and the Itô formula yield

$$
\begin{aligned}
\mathrm{d} \Theta_{t}^{\lambda, \mu}\left(X_{t}^{\nu}\right)= & \lambda \mathbf{u}_{t}^{\lambda, \mu}\left(X_{t}^{\nu}\right) \mathrm{d} t+\left(\left\{\nabla \Theta_{t}^{\lambda, \mu}\right\} \sigma_{t}\right)\left(X_{t}^{\nu}\right) \mathrm{d} W_{t} \\
& \quad+\left[\left\{\nabla \mathbf{u}_{t}^{\lambda, \mu}\right\}\left(b_{t}^{\nu}-b_{t}^{\mu}\right)+b_{t}^{\nu}-\bar{b}_{t}^{\mu}\right]\left(X_{t}^{\nu}\right) \mathrm{d} t \\
= & {\left[\lambda \mathbf{u}_{t}^{\lambda, \mu}+\left\{\nabla \Theta_{t}^{\lambda, \mu}\right\}\left(b_{t}^{\nu}-b_{t}^{\mu}\right)+\hat{b}_{t}^{\mu}\right]\left(X_{t}^{\nu}\right) \mathrm{d} t+\left(\left\{\nabla \Theta_{t}^{\lambda, \mu}\right\} \sigma_{t}\right)\left(X_{t}^{\nu}\right) \mathrm{d} W_{t} . }
\end{aligned}
$$

Combining this with (2.16) and applying the Itô formula, we see that $\eta_{t}:=\Theta_{t}^{\lambda, \mu}\left(X_{t}^{\mu}\right)-$ $\Theta_{t}^{\lambda, \mu}\left(X_{t}^{\nu}\right)$ satisfies

$$
\begin{aligned}
\mathrm{d}\left|\eta_{t}\right|^{2}= & 2\left\langle\eta_{t}, \lambda \mathbf{u}_{t}^{\lambda, \mu}\left(X_{t}^{\mu}\right)-\lambda \mathbf{u}_{t}^{\lambda, \mu}\left(X_{t}^{\nu}\right)+\hat{b}_{t}^{\mu}\left(X_{t}^{\mu}\right)-\hat{b}_{t}^{\mu}\left(X_{t}^{\nu}\right)\right\rangle \mathrm{d} t \\
& +2\left\langle\eta_{t},\left[\left(\left\{\nabla \Theta_{t}^{\lambda, \mu}\right\} \sigma_{t}\right)\left(X_{t}^{\mu}\right)-\left(\left\{\nabla \Theta_{t}^{\lambda, \mu}\right\} \sigma_{t}\right)\left(X_{t}^{\nu}\right)\right] \mathrm{d} W_{t}\right\rangle \\
& +\left\|\left(\left\{\nabla \Theta_{t}^{\lambda, \mu}\right\} \sigma_{t}\right)\left(X_{t}^{\mu}\right)-\left(\left\{\nabla \Theta_{t}^{\lambda, \mu}\right\} \sigma_{t}\right)\left(X_{t}^{\nu}\right)\right\|_{H S}^{2} \mathrm{~d} t \\
& -2\left\langle\eta_{t},\left[\left\{\nabla \Theta_{t}^{\lambda, \mu}\right\}\left(b_{t}^{\nu}-b_{t}^{\mu}\right)\right]\left(X_{t}^{\nu}\right)\right\rangle \mathrm{d} t .
\end{aligned}
$$

So, for any $m \geq 1$, it holds

$$
\begin{align*}
\mathrm{d}\left|\eta_{t}\right|^{2 m}= & 2 m\left|\eta_{t}\right|^{2(m-1)}\left\langle\eta_{t}, \lambda \mathbf{u}_{t}^{\lambda, \mu}\left(X_{t}^{\mu}\right)-\lambda \mathbf{u}_{t}^{\lambda, \mu}\left(X_{t}^{\nu}\right)+\hat{b}_{t}^{\mu}\left(X_{t}^{\mu}\right)-\hat{b}_{t}^{\mu}\left(X_{t}^{\nu}\right)\right\rangle \mathrm{d} t \\
& +2 m\left|\eta_{t}\right|^{2(m-1)}\left\langle\eta_{t},\left[\left(\left\{\nabla \Theta_{t}^{\lambda, \mu}\right\} \sigma_{t}\right)\left(X_{t}^{\mu}\right)-\left(\left\{\nabla \Theta_{t}^{\lambda, \mu}\right\} \sigma_{t}\right)\left(X_{t}^{\nu}\right)\right] \mathrm{d} W_{t}\right\rangle \\
& +m\left|\eta_{t}\right|^{2(m-1)}\left\|\left(\left\{\nabla \Theta_{t}^{\lambda, \mu}\right\} \sigma_{t}\right)\left(X_{t}^{\mu}\right)-\left(\left\{\nabla \Theta_{t}^{\lambda, \mu}\right\} \sigma_{t}\right)\left(X_{t}^{\nu}\right)\right\|_{H S}^{2} \mathrm{~d} t  \tag{2.17}\\
& +2 m(m-1)\left|\eta_{t}\right|^{2(m-2)}\left|\left[\left(\left\{\nabla \Theta_{t}^{\lambda, \mu}\right\} \sigma_{t}\right)\left(X_{t}^{\mu}\right)-\left(\left\{\nabla \Theta_{t}^{\lambda, \mu}\right\} \sigma_{t}\right)\left(X_{t}^{\nu}\right)\right]^{*} \eta_{t}\right|^{2} \mathrm{~d} t \\
& -2 m\left|\eta_{t}\right|^{2(m-1)}\left\langle\eta_{t},\left[\left\{\nabla \Theta_{t}^{\lambda, \mu}\right\}\left(b_{t}^{\nu}-b_{t}^{\mu}\right)\right]\left(X_{t}^{\nu}\right)\right\rangle \mathrm{d} t .
\end{align*}
$$

By (2.14) and (1.8), we may find a constant $c_{0}>0$ such that

$$
\begin{equation*}
\left|\eta_{t}\right|^{2(m-1)}\left|\eta_{t}\right| \cdot\left|\lambda \mathbf{u}_{t}^{\lambda, \mu}\left(X_{t}^{\mu}\right)-\lambda \mathbf{u}_{t}^{\lambda, \mu}\left(X_{t}^{\nu}\right)+\hat{b}_{t}^{\mu}\left(X_{t}^{\mu}\right)-\hat{b}_{t}^{\mu}\left(X_{t}^{\nu}\right)\right| \leq c_{0}\left|\eta_{t}\right|^{2 m} \tag{2.18}
\end{equation*}
$$

and

$$
\begin{align*}
& \left|\eta_{t}\right|^{2(m-1)}\left|\eta_{t}\right| \cdot\left|\left[\left\{\nabla \Theta_{t}^{\lambda, \mu}\right\}\left(b_{t}^{\nu}-b_{t}^{\mu}\right)\right]\left(X_{t}^{\nu}\right)\right| \\
& \leq K_{2}\left\|\nabla \Theta^{\lambda, \mu}\right\|_{\infty}\left|\eta_{t}\right|^{2(m-1)}\left|\eta_{t}\right|\left(\mathbb{W}_{\theta}\left(\mu_{t}, \nu_{t}\right)+\left\|\mu_{t}-\nu_{t}\right\|_{T V}\right)  \tag{2.19}\\
& \leq c_{0}\left(\left|\eta_{t}\right|^{2 m}+\mathbb{W}_{\theta}\left(\mu_{t}, \nu_{t}\right)^{2 m}+\left\|\mu_{t}-\nu_{t}\right\|_{T V}^{2 m}\right)
\end{align*}
$$

According to [13, (4.19)-(4.20)], we arrive at

$$
\begin{equation*}
\mathrm{d}\left|\eta_{t}\right|^{2 m} \leq c_{1}\left|\eta_{t}\right|^{2 m} \mathrm{~d} A_{t}+c_{1}\left(\mathbb{W}_{\theta}\left(\mu_{t}, \nu_{t}\right)^{2 m}+\left\|\mu_{t}-\nu_{t}\right\|_{T V}^{2 m}\right) \mathrm{d} t+\mathrm{d} M_{t} \tag{2.20}
\end{equation*}
$$

for some constant $c_{1}>0$, a local martingale $M_{t}$, and

$$
A_{t}:=\int_{0}^{t}\left\{1+\left(\mathscr{M}\left(\left\|\nabla^{2} \Theta_{s}^{\lambda, \mu}\right\|+\left\|\nabla \sigma_{s}\right\|\right)\left(X_{s}^{\mu}\right)+\mathscr{M}\left(\left\|\nabla^{2} \Theta_{s}^{\lambda, \mu}\right\|+\left\|\nabla \sigma_{s}\right\|\right)\left(X_{s}^{\nu}\right)\right)^{2}\right\} \mathrm{d} s
$$

Thanks to [24, Theorem 3.1], the Krylov estimate

$$
\begin{align*}
& \mathbb{E}\left[\int_{s}^{t}\left|f_{r}\right|\left(X_{r}^{\mu}\right) \mathrm{d} r \mid \mathscr{F}_{s}\right]+\mathbb{E}\left[\int_{s}^{t}\left|f_{r}\right|\left(X_{r}^{\nu}\right) \mathrm{d} r \mid \mathscr{F}_{s}\right] \\
& \leq C\left(\int_{s}^{t}\left(\int_{\mathbb{R}^{d}}\left|f_{r}(x)\right|^{p} \mathrm{~d} x\right)^{\frac{q}{p}} \mathrm{~d} r\right)^{\frac{1}{q}}, 0 \leq s<t \leq T . \tag{2.21}
\end{align*}
$$

holds. As shown in $\left[23\right.$, Lemma 3.5], (2.21), (2.5), (2.15) and $\left(A_{\sigma}\right)$ imply

$$
\sup _{t \in[0, T]} \mathbb{E} \mathrm{e}^{\delta A_{t}}=\mathbb{E} \mathrm{e}^{\delta A_{T}}<\infty, \quad \delta>0
$$

By (2.14) and the stochastic Gronwall lemma (see [23, Lemma 3.8]), (2.20) with $2 m>\theta$ implies

$$
\begin{aligned}
& \left\{\mathbb{W}_{\theta}\left(\Phi_{t}^{\gamma_{1}}(\mu), \Phi_{t}^{\gamma_{2}}(\nu)\right)\right\}^{2 m} \leq c_{2}\left(\mathbb{E}\left|\eta_{t}\right|^{\theta}\right)^{\frac{2 m}{\theta}} \\
& \leq c_{3}\left(\mathbb{E}\left|X_{0}^{\mu}-X_{0}^{\nu}\right|^{\theta}\right)^{\frac{2 m}{\theta}}+c_{3}\left(\mathbb{E}^{\frac{c_{1} \theta}{2 m-\theta}} A_{T}\right)^{\frac{2 m-\theta}{\theta}} \int_{0}^{t}\left(\mathbb{W}_{\theta}\left(\mu_{s}, \nu_{s}\right)^{2 m}+\left\|\mu_{s}-\nu_{s}\right\|_{T V}^{2 m}\right) \mathrm{d} s
\end{aligned}
$$

holds for all $t \in[0, T]$ and some constants $c_{2}, c_{3}>0$. Therefore, (2.12) holds for some constant $C>0$ and the proof is thus finished.

## 3 Proof of Theorem 1.1

Assume $\left(A_{\sigma}\right)$. According to [25, Theorem 1.3], for any $\mu$. $\in \mathscr{B}([0, T] ; \mathscr{P})$, each of $\left(A_{b}\right)$ and $\left(A_{b}^{\prime}\right)$ implies the strong existence and uniqueness up to life time of the $\operatorname{SDE}(2.1)$. Moreover, it is standard that in both cases a solution of (2.1) is non-explosive. So, by Lemma 2.1, the strong well-posedness of (1.1) implies the weak well-posedness. Therefore, in the following we need only cosnider the strong solution.

To prove the strong well-posedness of (1.1), it suffices to find a constant $t_{0} \in(0, T]$ independent of $X_{0}$ such that in each of these two cases the SDE (1.1) has strong wellposedness up to time $t_{0}$. Indeed, once this is confirmed, by considering the SDE from time $t_{0}$ we prove the same property up to time $\left(2 t_{0}\right) \wedge T$. Repeating the procedure finite many times we derive the strong well-posedness.

Below we prove assertions (1) and (2) for strong solutions respectively.
(a) Let $\left(A_{b}^{\prime}\right)$ hold. Take $t_{0}=\min \left\{T, \frac{1}{K_{1} K_{3}^{2}}\right\}$ and consider the space $E_{t_{0}}:=\{\mu \in$ $\left.\mathscr{B}\left(\left[0, t_{0}\right] ; \mathscr{P}\right): \mu_{0}=\gamma\right\}$ equipped with the complete metric

$$
\rho(\nu, \mu):=\sup _{t \in\left[0, t_{0}\right]}\left\|\nu_{t}-\mu_{t}\right\|_{T V} .
$$

Then (2.7) implies that $\Phi^{\gamma}$ is a strictly contractive map on $E_{t_{0}}$, so that it has a unique fixed point, i.e. the equation

$$
\begin{equation*}
\Phi_{t}^{\gamma}(\mu)=\mu_{t}, \quad t \in\left[0, t_{0}\right] \tag{3.1}
\end{equation*}
$$

has a unique solution $\mu \in E_{t_{0}}$. By (3.1) and the definition of $\Phi^{\gamma}$ we see that the unique solution of (2.1) is a strong solution of (1.1). On the other hand, $\mu_{t}:=\mathscr{L}_{X_{t}}$ for any strong solution to (1.1) is a solution to (3.1), hence the uniqueness of (3.1) implies that of (1.1).

To prove (1.12), let $\mu_{t}=P_{t}^{*} \mu_{0}$ and $\nu_{t}=P_{t}^{*} \nu_{0}$. We have $P_{t}^{*} \mu_{0}=\Phi_{t}^{\mu_{0}}(\mu)$ and $P_{t}^{*} \nu_{0}=$ $\Phi_{t}^{\nu_{0}}(\nu)$. So, (2.7) with $\gamma=\mu_{0}$ implies

$$
\begin{equation*}
\left\|P_{t}^{*} \mu_{0}-\Phi_{t}^{\mu_{0}}(\nu)\right\|_{T V}^{2} \leq \frac{K_{1} K_{3}^{2}}{4} \int_{0}^{t}\left\|P_{s}^{*} \mu_{0}-P_{s}^{*} \nu_{0}\right\|_{T V}^{2} \mathrm{~d} s, \quad t \in[0, T] . \tag{3.2}
\end{equation*}
$$

On the other hand, by the Markov property for the solution to (2.6) with $\nu$ replacing $\mu$, we have

$$
\Phi_{t}^{\gamma}(\nu)=\int_{\mathbb{R}^{d}} \Phi_{t}^{\delta_{x}}(\nu) \gamma(\mathrm{d} x), \quad \gamma \in \mathscr{P} .
$$

Combining this with $P_{t}^{*} \nu_{0}=\Phi_{t}^{\nu_{0}}(\nu)$, we obtain

$$
\begin{aligned}
\left|\left\{\Phi_{t}^{\mu_{0}}(\nu)\right\}(A)-\left\{P_{t}^{*} \nu_{0}\right\}(A)\right| & =\left|\int_{\mathbb{R}^{d}}\left\{\Phi_{t}^{\delta_{x}}(\nu)\right\}(A)\left(\mu_{0}-\nu_{0}\right)(\mathrm{d} x)\right| \\
& \leq\left\|\mu_{0}-\nu_{0}\right\|_{T V}, \quad A \in \mathscr{B}\left(\mathbb{R}^{d}\right) .
\end{aligned}
$$

Hence,

$$
\begin{equation*}
\left\|\Phi_{t}^{\mu_{0}}(\nu)-P_{t}^{*} \nu_{0}\right\|_{T V} \leq\left\|\mu_{0}-\nu_{0}\right\|_{T V}, \quad t \in[0, T] . \tag{3.3}
\end{equation*}
$$

This together with (3.2) yields

$$
\begin{aligned}
& \left\|P_{t}^{*} \mu_{0}-P_{t}^{*} \nu_{0}\right\|_{T V}^{2} \leq 2\left\|P_{t}^{*} \mu_{0}-\Phi_{t}^{\mu_{0}}(\nu)\right\|_{T V}^{2}+2\left\|\Phi_{t}^{\mu_{0}}(\nu)-P_{t}^{*} \nu_{0}\right\|_{T V}^{2} \\
& \leq 2\left\|\mu_{0}-\nu_{0}\right\|_{T V}^{2}+\frac{K_{1} K_{3}^{2}}{2} \int_{0}^{t}\left\|P_{s}^{*} \mu_{0}-P_{s}^{*} \nu_{0}\right\|_{T V}^{2} \mathrm{~d} s, \quad t \in[0, T]
\end{aligned}
$$

By Gronwall's lemma, this implies (1.12).
(b) Let $\left(A_{b}\right)$ hold and let $\gamma=\mathscr{L}_{X_{0}} \in \mathscr{P}_{\theta}$. For any $\mu, \nu \in C\left([0, T], \mathscr{P}_{\theta}\right)$, (1.8) implies (2.11). By (2.11), (1.8) and (2.8) with $\gamma_{1}=\gamma_{2}=\gamma$, we find a constant $C>0$ such that

$$
\begin{align*}
& \left\{\left\|\Phi_{t}^{\gamma}(\mu)-\Phi_{t}^{\gamma}(\nu)\right\|_{T V}+\mathbb{W}_{\theta}\left(\Phi_{t}^{\gamma}(\mu), \Phi_{t}^{\gamma}(\nu)\right)\right\}^{2 m} \\
& \leq C \int_{0}^{t}\left\{\left\|\mu_{s}-\nu_{s}\right\|_{T V}+\mathbb{W}_{\theta}\left(\mu_{s}, \nu_{s}\right)\right\}^{2 m} \mathrm{~d} s, \quad t \in[0, T], \gamma \in \mathscr{P}_{\theta} \tag{3.4}
\end{align*}
$$

Let $t_{0}=\frac{1}{2 C}$. We consider the space $\tilde{E}_{t_{0}}:=\left\{\mu \in C\left(\left[0, t_{0}\right] ; \mathscr{P}_{\theta}\right): \mu_{0}=\gamma\right\}$ equipped with the complete metric

$$
\tilde{\rho}(\nu, \mu):=\sup _{t \in\left[0, t_{0}\right]}\left\{\left\|\nu_{t}-\mu_{t}\right\|_{T V}+\mathbb{W}_{\theta}\left(\nu_{t}, \mu_{t}\right)\right\} .
$$

Then $\Phi^{\gamma}$ is strictly contractive in $\tilde{E}_{t_{0}}$, so that the same argument in (a) proves the strong well-posedness of (1.1) with $\mathscr{L}_{X_{0}}=\gamma$ up to time $t_{0}$.

Let $\mu_{t}$ and $\nu_{t}$ be in (a). By (3.4) with $\gamma=\mu_{0}$ we obtain

$$
\begin{align*}
& \left\{\left\|P_{t}^{*} \mu_{0}-\Phi_{t}^{\mu_{0}}(\nu)\right\|_{T V}+\mathbb{W}_{\theta}\left(P_{t}^{*} \mu_{0}, \Phi_{t}^{\mu_{0}}(\nu)\right)\right\}^{2 m} \\
& \leq C \int_{0}^{t}\left\{\left\|P_{s}^{*} \mu_{0}-P_{s}^{*} \nu_{0}\right\|_{T V}+\mathbb{W}_{\theta}\left(P_{s}^{*} \mu_{0}, P_{s}^{*} \nu_{0}\right)\right\}^{2 m} \mathrm{~d} s, \quad t \in[0, T] . \tag{3.5}
\end{align*}
$$

Next, taking $\gamma_{1}=\nu_{0}, \gamma_{2}=\mu_{0}$ and $\mu=\nu$ in (2.8), we derive

$$
\left\{\mathbb{W}_{\theta}\left(P_{t}^{*} \nu_{0}, \Phi_{t}^{\mu_{0}}(\nu)\right)\right\}^{2 m} \leq C\left\{\mathbb{W}_{\theta}\left(\mu_{0}, \nu_{0}\right)\right\}^{2 m}
$$

Combining this with (3.3) and (3.5), we find a constant $C^{\prime}>0$ such that

$$
\begin{aligned}
& \left\{\left\|P_{t}^{*} \mu_{0}-P_{t}^{*} \nu_{0}\right\|_{T V}+\mathbb{W}_{\theta}\left(P_{t}^{*} \mu_{0}, P_{t}^{*} \nu_{0}\right)\right\}^{2 m} \\
& \leq C^{\prime}\left\{\left\|\mu_{0}-\nu_{0}\right\|_{T V}+\mathbb{W}_{\theta}\left(\mu_{0}, \nu_{0}\right)\right\}^{2 m} \\
& +C^{\prime} \int_{0}^{t}\left\{\left\|P_{s}^{*} \mu_{0}-P_{s}^{*} \nu_{0}\right\|_{T V}+\mathbb{W}_{\theta}\left(P_{s}^{*} \mu_{0}, P_{s}^{*} \nu_{0}\right)\right\}^{2 m} \mathrm{~d} s, \quad t \in[0, T]
\end{aligned}
$$

By Gronwall's lemma, this implies (1.13) for some constant $c>0$.

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